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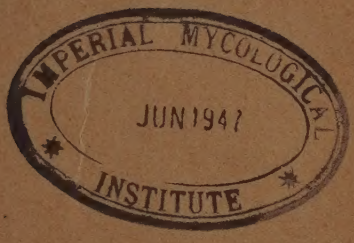

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MATHÉMATIQUES

SOMMATION DES SUITES BORNÉES PAR LES MÉTHODES LINÉAIRES RÉGULIÈRES

Par A. L. BROUDNO

(Présenté par A. N. Kolmogoroff, de l'Académie, le 28.X 1943)

La sommation des suites bornées par les méthodes linéaires régulières* jouit de certaines propriétés qui, en général, n'ont plus lieu quand on considère les sommations des suites arbitraires.

Par exemple: si une matrice B** de Toeplitz réalise la sommation de toutes les suites bornées, qui sont sommables par la matrice A, elle donne pour chacune de ces suites la même limite que celle qui est fournie par A (théorème 4). Cela permet de déterminer une méthode linéaire régulière non pas comme matrice de Toeplitz, mais directement au moyen de l'ensemble des suites bornées qui sont sommables par cette méthode.

Le but de la présente Note est d'étudier la situation mutuelle des ensembles de cette nature et de trouver quel lien il y a entre leur structure et les nombres que l'on fait correspondre en qualité de limites généralisées aux suites bornées qui appartiennent à ces ensembles.

Introduisons d'abord les définitions suivantes:

I. On appelle norme d'une matrice de Toeplitz $A=(a_{ik})$ la valeur

$$\|A\| = \sup \sum_{k=1}^{\infty} |a_{ik}|$$

Il suit de la définition d'une matrice de Toeplitz que sa norme est toujours unie.

II. On appelle champ borné d'une matrice A de Toeplitz l'ensemble $\mathfrak{Q}(A)$ de toutes les suites bornées $\{S_n\}$ qui sont sommables par la matrice A. On appellera «champ borné» tout court un ensemble de suites qui est le champ borné d'une certaine matrice de Toeplitz.

III. On appelle norme d'un champ $\mathfrak{Q}$ (notation: $\|\mathfrak{Q}\|$) la borne inférieure des normes de toutes les matrices de Toeplitz pour lesquelles $\mathfrak{Q}$ est le champ borné.

On appelle champ trivial l'ensemble de toutes les suites qui convergent vers des limites finies. $E_1 \subset E_2$ signifie que E_1 est un ensemble appartenant à E_2, mais ne coïncidant pas avec E_2. $E_1 \subseteq E_2$ signifie que E_1 appartient à E_2 et peut coïncider avec E_2. Une suite d'ensembles $E_1, E_2, \ldots, E_n, \ldots$ est

* Une méthode linéaire est appelée régulière, si la convergence d'une suite vers une valeur finie entraîne sa sommabilité par la méthode donnée et vers la même valeur. Dans la présente Note nous ne considérons que les méthodes linéaires définies par des matrices.

** Il est connu qu'une méthode linéaire, définie par une matrice, est régulière dans le cas et dans ce cas seulement où la matrice considérée est une matrice de Toeplitz (1).

appelée croissante (décroissante), si $E_n \subset E_{n1}$ $(E_n \supset E_{n+1})$ pour toutes les valeurs de n.

On peut démontrer les théorèmes suivants.

Théorème 1. *Pour chaque nombre $N \geqslant 1$ il existe un champ borné Ω tel que $\|\Omega\| = N$.*

Théorème 2. *Si le champ borné Ω_1 est contenu dans le champ borné Ω_2, on a $\|\Omega_1\| \leqslant \|\Omega_2\|$.*

Théorème 3. *Il existe une suite croissante de champs bornés dont les normes croissent indéfiniment.*

Il en résulte que la somme au sens de la théorie des ensembles d'une suite croissante de champs bornés peut jouir de cette propriété qu'elle n'appartient à aucun champ borné.

On sait que: quels que soient la suite divergente $S_1, S_2, \ldots, S_n, \ldots$ et le nombre S, il existe une matrice de Toeplitz qui réalise la sommation de la suite $\{S_n\}$ et lui attribue comme limite le nombre S(2).

Cependant on a le

Théorème 4. *Si une matrice B de Toeplitz réalise la sommation de toutes les suites bornées sommables par une matrice A de Toeplitz, alors elle leur attribue les mêmes limites que la matrice A.*

En d'autres termes, un champ borné Ω détermine d'une manière unique les limites qu'il faut attribuer à toutes les suites $\{S_n\} \in \Omega$ sommables par des méthodes linéaires régulières, pourvu que chacune de ces méthodes réalise la sommation de toutes les suites $\{S_n\} \in \Omega$.

Cela étant, on peut introduire la notion de limite (unique) qu'un champ borné Ω attribue à une suite $\{S_n\} \in \Omega$, à savoir: nous dirons qu'une suite $\{S_n\} \in \Omega$ est sommable par le champ Ω et a pour limite S, s'il existe au moins une matrice de Toeplitz qui transforme toutes les suites appartenant à Ω en suites convergentes et qui transforme la suite $\{S_n\}$ en une suite convergente qui a pour limite S.

Ainsi on a le

Théorème 4'. *Un champ borné détermine d'une manière unique les limites qu'il attribue aux suites appartenant à ce champ.*

Et plus généralement:

Théorème 4''. *Si un champ borné Ω_1 est contenu dans un champ borné Ω_2, chaque suite $\{S_n\}$ appartenant au champ borné Ω_1 est sommable par le champ Ω_2, et la limite attribuée par Ω_2 coïncide avec celle qui est attribuée par Ω_1.*

On sait que les suites bornées $\{S_n\}$ ayant pour norme

$$\|\{S_n\}\| = \sup_n |S_n|$$

forment un espace de Banach. Les théorèmes précédents permettent de démontrer que dans cet espace un ensemble dénombrable et de même un ensemble séparable de suites bornées n'est jamais un champ borné non-trivial.

Cette dernière proposition a été démontrée précédemment par Klebanoff (elle se trouve dans sa thèse qui n'a pas été publiée).

Comme conséquence du théorème 4'' on obtient encore le

Théorème 5. *Si deux champs bornés contiennent une certaine suite, mais les limites qu'ils lui attribuent sont différentes, alors leur somme (au sens de la théorie des ensembles) n'est contenue dans aucun champ borné.*

Il suit de ce théorème qu'il existe des champs bornés Ω_1 et Ω_2 dont la somme au sens de la théorie des ensembles n'est contenue dans aucun champ borné.

Théorème 6. *Si Ω_1 et Ω_2 sont des champs bornés, il existe un champ borné Ω contenu dans la partie commune de Ω_1 et Ω_2 et qui contient chaque*

champ borné $\omega \subseteq \Omega_1 \cdot \Omega_2$. *Ce champ borné* Ω *coïncide avec l'ensemble de toutes les suites telles que le champ* Ω_1 *leur attribue la même limite que le champ* Ω_2.

Théorème 7. *La partie commune d'une suite décroissante de champs bornés est un champ borné et contient au moins une suite bornée divergente.*

Théorème 8. *Si un champ borné* Ω *est contenu dans la somme (au sens de la théorie des ensembles) d'un nombre fini ou bien d'une infinité dénombrable de champs bornés* $\Omega_1, \Omega_2, \ldots, \Omega_n, \ldots$, *alors il est contenu dans l'un d'eux.*

Corollaire. La somme (au sens de la théorie des ensembles) d'une suite croissante de champs bornés $\Omega_1 \subset \Omega_2 \subset \ldots \subset \Omega_n \subset \ldots$ n'est jamais un champ borné.

Théorème 9. *Si* Ω_0 *et* Ω_1 *sont des champs bornés tels que* $\Omega_0 \subset \Omega_1$, *alors il existe un champ borné* $\Omega_{1/2}$ *tel que* $\Omega_0 \subset \Omega_{1/2} \subset \Omega_1$.

Il suit des théorèmes 7 et 9 que l'on a

Théorème 10. *Si* Ω_a *et* Ω_b *sont des champs bornés tels que* $\Omega_a \subset \Omega_b$, *alors à chaque nombre réel* ν, $0 \leqslant \nu \leqslant 1$, *on peut faire correspondre un champ borné* Ω_ν *tel que*

1) $\Omega_a \subseteq \Omega_0 \subset \Omega_1 = \Omega_b$;

2) *de* $0 \leqslant \nu < \mu \leqslant 1$ *il résulte* $\Omega_\nu \subset \Omega_\mu$;

3) *si* $0 \leqslant \nu \leqslant \mu_n \leqslant 1$ *pour tous les* $n = 1; 2, \ldots$ *et si* $\lim_{n\to\infty} \mu_n = \nu$, *alors* $\Omega_\nu = \lim_{n\to\infty} \Omega_{\mu_n}$.

Remarquons que dans la condition 1) la relation $\Omega_a \subseteq \Omega_0$ ne peut pas être remplacée par l'égalité $\Omega_a = \Omega_0$, du moins dans le cas où Ω_a est le champ trivial (voir le théorème 7). D'ailleurs, le théorème 8 montre que dans la condition 3) on ne peut pas exclure l'hypothèse $\mu_n \geqslant \nu$.

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Moscou.

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MATHÉMATIQUES

SUR LA PUISSANCE CONSTRUCTIVE D'UN COMPLEXE *E* SUR LE PLAN DE LOBATCHEVSKI

Par N. M. NESTOROVITCH

(Présenté par A. N. Kolmogoroff, de l'Académie, le 28.X.1943)

§ 1. Nous avons démontré [1–2] que tous les problèmes d'ordre constructif sur le plan de Lobatchevski qui conduisent à l'équidistante peuvent être résolus par le complexe classique *E* (compas—règle) sans utiliser un instrument spécial—l'hypercompas. Suivant une proposition du professeur W. F. Kagan, j'ai fait une étude analogue par rapport à l'horicycle. J'ai trouvé que le complexe *E* suffit encore pour résoudre les problèmes qui conduisent à l'horicycle comme lieu géométrique des points; ainsi, comme dans le cas de l'équidistante, il n'est pas nécessaire d'utiliser un instrument nouveau—un horicompas—pour résoudre les problèmes qui conduisent à l'horicycle *.

§ 2. Pour établir ce fait, démontrons d'abord que toutes les méthodes pour définir un horicycle [voir § 4, méthodes b), c), d)] peuvent être réduites à sa détermination au moyen de a): un axe dirigé $\overrightarrow{AA'}$ (un centre à l'infini) et un point A qui est situé sur cet axe et appartient à l'horicycle. Cette manière de définir un horicycle correspond à la manière habituelle de définir une circonférence par la position de son centre O et la valeur de son rayon r, et de même de définir une équidistante par son axe OQ et sa distance h.

Convenons de désigner par le symbole $\smile \| \overrightarrow{AA'} \| (A)$ un horicycle défini de cette manière.

§ 3. Remarquons d'abord que si l'on connaît l'axe dirigé $\overrightarrow{AA'}$ de l'horicycle et un point A sur cet axe, on peut construire par le complexe E un nombre fini quelconque de points de l'horicycle. En effet, menons un rayon AD qui forme avec l'axe $\overrightarrow{AA'}$ un angle aigu ξ arbitraire, alors il est facile de construire $AD = \Delta(\xi)$ [3]. Le segment $x = AB = 2\Delta(\xi)$ définit sur le rayon AD un point B appartenant à l'horicycle. Il est évident que le point B de l'horicycle est défini ou bien par l'angle ξ entre la corde AB et l'axe $\overrightarrow{AA'}$, ou bien par la longueur de la corde x, parce que $\xi = \Pi\left(\frac{x}{2}\right)$.

§ 4. Supposons maintenant que l'horicycle est défini par d'autres méthodes, à savoir: b) par l'axe $\overrightarrow{AA'}$ et un point P hors de cet axe et à distance

* En ce qui concerne la construction d'un horicompas, on peut se présenter qu'il est réalisé en forme d'un arc matériel d'un horicycle AC durement lié avec deux points A et B de l'axe de l'horicycle AB, et d'ailleurs l'un de ces points, par exemple A, doit appartenir à l'arc de l'horicycle. Si l'on considère d'ailleurs l'axe de l'horicycle comme réalisé matériellement en forme d'une règle AB, alors l'horicompas aura la forme d'un équerre dont un côté est droit et l'autre est une courbe (arc de l'horicycle).

h de $\overrightarrow{AA'}$. Menons $PC' \parallel \overrightarrow{AA'}$ et trouvons sur $\overrightarrow{AA'}$ un point A dont la distance à PC' est égale à h(2). Le point A de l'axe $\overrightarrow{AA'}$ appartiendra à l'horicycle.

c) Supposons ensuite que l'horicycle est défini par trois points M, N, P et que d'ailleurs $\angle MNP = \Pi\left(\frac{MN}{2}\right) + \Pi\left(\frac{NP}{2}\right)$. En menant par les milieux A et B des cordes MN et NP les médiatrices AA' et BB', déterminons l'axe dirigé $\overrightarrow{AA'}$ de l'horicycle. En possédant encore un point hors de l'axe, par exemple M, nous reviendrons à la méthode précédente b).

d) Enfin, si l'horicycle est donné d'une manière graphique par un segment de l'arc, alors en choisissant trois points sur cet arc, on réduit cette méthode à la méthode c).

§ 5. L'application de l'horicompas dans les problèmes constructifs se réduit à la résolution de l'un des 5 problèmes suivants: la recherche du point d'intersection: 1° d'une droite et d'un horicycle, 2° d'une circonférence et d'un horicycle, 3° d'une équidistante et d'un horicycle, 4° de deux horicycles, et 5° le tracement d'un horicycle.

En ce qui concerne le problème 5°, il est vrai que pour tracer un horicycle comme un lieu géométrique continu, il faut avoir un instrument spécial, mais un nombre quelconque de points isolés peut être obtenu par le complexe E (§ 3). Quant même les points cités dans les problèmes 1° — 4° peuvent être déterminés (voir les §§ 6 — 10) au moyen de constructions du premier et du second ordre (dans tous les cas ces constructions sont de première classe), donc, d'après un théorème de D. Mordouchaï-Boltovsky (4) les problèmes 1° — 4° sont résolubles par le complexe E.

§ 6. Problème 1°. Cas α), où la sécante CB rencontre l'axe $\overrightarrow{AA'}$ de l'horicycle au point C. Le segment x, qui détermine le point B cherché, est lié avec les paramètres b et ν, qui définissent la situation commune de la sécante CB et de l'axe AA', par l'équation

$$2(\operatorname{sh} b \pm \operatorname{ch} b)\operatorname{sh}^2\frac{x}{2} - 2\operatorname{ctg}\nu \operatorname{sh}\frac{x}{2} + \operatorname{sh} b = 0^* \qquad (1)$$

Ici ν est l'angle $(CB, \overrightarrow{AA'})$, $b = CA$, $x = AB$ est longueur de la corde $\frown\overrightarrow{AA'}(A)$.

Cas β), où la sécante BB' est surparallèle à l'axe $\overrightarrow{AA'}$. Pour la longueur x de la corde on obtient l'équation

$$2(\operatorname{ch} b \pm \operatorname{sh} b)\operatorname{sh}^2\frac{x}{2} - 2\operatorname{ctg} d \operatorname{sh}\frac{x}{2} + \operatorname{ch} b = 0 \qquad (2)$$

Ici $d = DC$ est la longueur de la perpendiculaire commune de BB' et $\overrightarrow{AA'}$; b et x ont les mêmes valeurs que précédemment.

Cas γ), où la sécante $CC' \parallel AA'$; le problème se réduit à la construction du § 4, b).

§ 7. Le problème 2° se réduit à l'équation

$$2\operatorname{ch} a(\operatorname{ch} d \pm \operatorname{sh} d)\operatorname{sh}^2\frac{x}{2} - 2\operatorname{sh} a \operatorname{sh}\frac{x}{2} + \operatorname{ch} a \operatorname{ch} d - \operatorname{ch} r = 0 \qquad (3)$$

Ici r est le rayon du cercle donné, et $a = OP$ est la distance de son centre O de l'axe $\overrightarrow{AA'}$; $d = PA$, x est la longueur de la corde AM, M est le point cherché.

---

* Le signe double dans l'équation (1) et les suivantes correspond aux cas différents de la situation mutuelle des lignes qui se coupent.

§ 8. Problème 3°. Cas α), où $\overrightarrow{AA'}$ et l'axe OQ de l'équidistante forment un angle ν. Pour déterminer x nous avons l'équation

$$2(\operatorname{sh} a - \operatorname{ch} a \cos\lambda)\operatorname{sh}^2\frac{x}{2} \pm 2\operatorname{ch} a \sin\lambda \operatorname{sh}\frac{x}{2} + \operatorname{sh} a \mp \operatorname{sh} h = 0 \qquad (4)$$

Ici $a = AP$ est la distance de A à OQ, $\lambda = \angle PAA'$, h est la distance de l'équidistante donnée, M est le point cherché, et $AM = x$.

Cas β), où $\overrightarrow{AA'}$ et OQ sont surparallèles; leur perpendiculaire commune $ED = \delta$.

Pour x on obtient l'équation

$$2(\operatorname{sh} a - \operatorname{ch} a \cos\beta)\operatorname{sh}^2\frac{x}{2} \pm 2\operatorname{ch} a \sin\beta \operatorname{sh}\frac{x}{2} + \operatorname{sh} a \mp \operatorname{sh} h = 0 \qquad (5)$$

Ici a et h ont les valeurs précédentes; $\beta = \angle PAA'$.

Cas γ), où $\overrightarrow{AA'} \parallel OQ$; x est déterminé par l'équation

$$\operatorname{sh}\frac{x}{2} = \pm\frac{\operatorname{sh} h - \operatorname{sh} a}{2} \qquad (6)$$

§ 9. Le problème 4° se réduit à l'équation

$$2[(\operatorname{ch} b - \operatorname{sh} b)(\operatorname{ch} c - \operatorname{sh} c \cos\mu) + \operatorname{sh} a - \operatorname{ch} a]\operatorname{sh}^2\frac{x}{2} -$$

$$-2(\operatorname{ch} b - \operatorname{sh} b)\operatorname{ch} c \sin\mu \operatorname{sh}\frac{x}{2} + \operatorname{ch} c(\operatorname{ch} b - \operatorname{sh} b) - \operatorname{ch} a + \operatorname{sh} b = 0 \qquad (7)$$

si les axes des horicycles donnés $\smile\overrightarrow{AA'}(A)$ et $\smile\overrightarrow{BB'}(B)$ se coupent, et à l'équation

$$2[(\operatorname{ch} b - \operatorname{sh} b)(\operatorname{ch} c - \operatorname{sh} c \cos\mu) + \operatorname{sh} a - \operatorname{ch} a]\operatorname{sh}^2\frac{x}{2} -$$

$$-2\operatorname{sh} c \sin\mu(\operatorname{ch} b - \operatorname{sh} b)\operatorname{sh}\frac{x}{2} + \operatorname{ch} c(\operatorname{ch} b - \operatorname{sh} b) + \operatorname{sh} a - \operatorname{ch} b = 0 \qquad (8)$$

si les axes sont surparallèles.

Ici a, b sont les distances des points A et B du point d'intersection C (réel) ou bien DE (idéal) des axes $\overrightarrow{AA'}$ et $\overrightarrow{BB'}$; $c = AB$; $\mu = \angle BAA'$.

§ 10. En considérant ces résultats en même temps que les résultats des travaux précédents [1] et [2], on arrive à la conclusion que le complexe E suffit pour résoudre tous les problèmes du second ordre sur le plan de Lobatchevski indépendamment du fait laquelle des trois circonférences de l'espace de Lobatchevski prend part dans la construction du problème considéré. Ainsi le rôle de l'horicompas et de l'hypercompas dans les constructions dans le plan de Lobatchevski est à peu près le même que le rôle de l'équerre et de la règle à deux bords dans les constructions sur le plan E; ils ne doivent pas nécessairement prendre part dans la solution des problèmes du second degré, mais ils simplifient les constructions. Il convient de remarquer ici que le hypercompas se réduit dans l'espace E en règle à deux bords, l'horicompas — en équerre (voir la description de la construction d'un horicompas dans le § 1).

Université Molotoff.
Rostov/Don.

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MÉCANIQUE

LE MOUVEMENT D'UN AVION NEUTRE PILOTÉ AUTOMATIQUEMENT ET LA THÉORIE DES TRANSFORMATIONS PONCTUELLES DES SURFACES

Par A. ANDRONOW et N. BAUTIN

(Présenté par L. I. Mandelstam, de l'Académie, le 8.II.1944)

Nous étudions dans cette Note, en faisant les simplifications usuelles, la stabilité de route d'un avion neutre (c'est-à-dire sans couple redresseur) muni d'un auto-pilote idéal commandant la vitesse angulaire du gouvernail de direction à l'aide d'un servo-moteur à vitesse constante *.

La théorie des transformations ponctuelles des surfaces permet de résoudre ce problème non-linéaire ** sans aucune hypothèse quant à la petitesse de telle ou telle grandeur figurant dans l'équation du mouvement.

1. Les équations du mouvement du système étudié peuvent être mises sous la forme ([1], [2], [3])

$$\ddot{\varphi}+M\dot{\varphi}=-N\eta;\quad \ddot{\eta}=F;\quad \psi=\varphi+\beta\dot{\varphi}-\frac{1}{a}\eta \tag{1}$$

où

$$F=K \text{ pour } \psi>0,\ F=-K \text{ pour } \psi<0,\ F=a(\dot{\varphi}+\beta\ddot{\varphi}) \text{ pour } \psi=0.$$

φ est l'angle de lacet de l'avion; η, le déplacement angulaire du gouvernail; ψ, l'argument du servo-moteur commandant le gouvernail; β, le coefficient dit d'amortissement artificiel; $\frac{1}{a}$, le coefficient dit de rétroaction; M et N, K, des constantes positives caractérisant respectivement l'amortissement intrinsèque de l'avion et le gouvernail.

Introduisant

$$t'=Mt,\ z=\frac{M^3}{NK}\frac{d^2\varphi}{dt'^2},\ u=\frac{M^3}{NK}\left(\frac{d\varphi}{dt'}+\frac{d^2\varphi}{dt'^2}\right);\ A=M\beta,\ B=\frac{M^2}{Na}$$

et supposant $A>0$, $B>0$, on obtient le système

$$\left.\begin{array}{c}
\frac{dz}{dt'}+z=f(u,\ z,\ \psi);\ \frac{du}{dt'}=f(u,\ z,\ \psi);\ \frac{d\psi}{dt'}=g(u,\ z,\ \psi)\\
f=-1,\ g=u+(A-1)z-B \text{ si } \psi>0, \text{ ou } \psi=0,\\
u+(A-1)z-B>0\\
\text{ou } \psi=0,\ u+(A-1)z-B=0,\ u>A+B;\\
f=+1,\ g=u+(A-1)z+B \text{ si } \psi<0, \text{ ou } \psi=0,\ u+(A-1)z+B<0\\
\text{ou } \psi=0,\ u+(A-1)z+B=0,\quad u<-(A+B);\\
f=-\frac{1}{B}[(A-1)z+u],\ g\equiv 0 \text{ si } \psi=0,\ -B<u+(A-1)z<B
\end{array}\right\} \tag{2}$$

* La théorie exposée ici s'applique, en variant convenablement le sens mécanique des constantes, au cas idéal de stabilisation anti-roulis d'un avion (dont les ailes ne sont pas en V) par un auto-pilote commandant la vitesse angulaire des ailerons à l'aide d'un servo-moteur à vitesse constante.

** Le problème non-linéaire de l'avion piloté automatiquement a été l'objet d'un nombre considérable de travaux récents ([2-9]).

Etudions les solutions du système (2) déterminées par les conditions initiales $t'=0$, $u=u_0$, $z=z_0$, $\psi=0$, et la condition de continuité des fonctions $u(t')$, $z(t')$, $\psi(t')$ aux points de discontinuité des seconds membres f et g.

2. Le système (6) possède l'unique état d'équilibre $u=0$, $z=0$, $\psi=0$ appartenant à la «lame» $-B<u+(A-1)z<B$ $(\psi=0)$ qui forme en quelque sorte un élément d'attraction pour les trajectoires du système. La «lame» est divisée en trajectoires de telle façon que l'état d'équilibre est un noeud stable lorsque $(A+B)^2<4B$ et un foyer stable si $(A+B)^2<4B$.

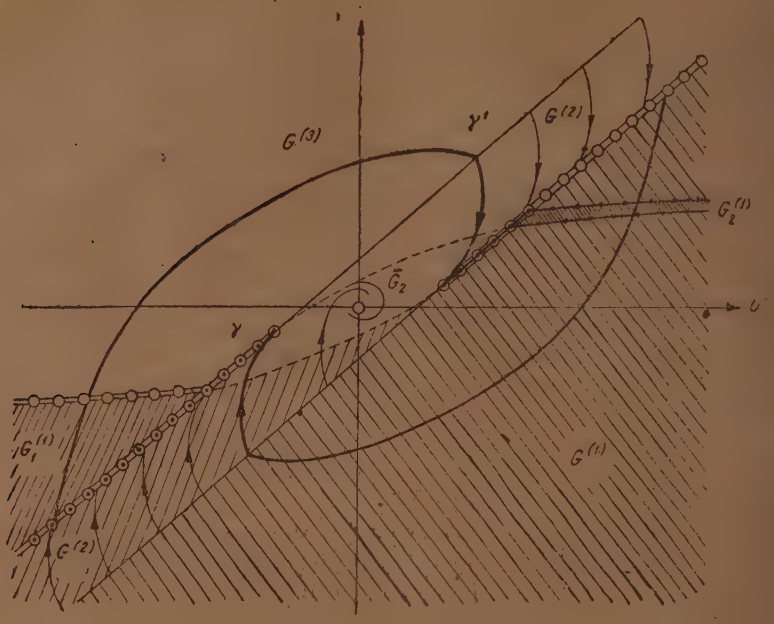

Fig. 1

3. Les équations (2) déterminent une transformation du plan $u, z\,(\psi=0)$ en lui-même. Partageons le plan $u, z\,(\psi=0)$ en quatre morceaux (fig. 1):

$G^{(1)}$, où soit $u+(A-1)z-B>0$, soit $u+(A-1)z-B=0$, $u>A+B$;

$G^{(3)}$, où soit $u+(A-1)z+B<0$, soit $u+(A-1)z+B=0$, $u<-(A+B)$;

$\bar{G}^{(2)}$ contenant l'origine et limité par les segments des droites $u+(A-1)z\pm\pm B=0$ et des courbes γ et γ'*; ces segments font partie de $\bar{G}^{(2)}$;

$G^{(2)}$, la partie du plan, contenant les points n'appartenant pas à $G^{(1)}$, $G^{(3)}$ et $\bar{G}^{(2)}$.

Si le point initial $u=u_0$, $z=z_0$ appartient à $\bar{G}^{(2)}$, en vertu de (2) il rejoint l'origine sans quitter le plan $\psi=0$. La transformation du point u_0, z_0 appartenant à $\bar{G}^{(2)}$ en le point (0,0) sera dite transformation $\bar{E}$. Si le point initial u_0, z_0 appartient à $G^{(2)}$, en vertu de (2) il rejoint sans quitter le plan $\psi=0$ le point $u_1=-(A-1)z_1\pm B$, z_1 appartenant à $G^{(3)}$ ou $G^{(1)}$. Cette transformation du point u_0, z_0 sera dite transformation E. Si le point initial u_0, $z_0\,(t'=0)$ appartient à $G^{(1)}$ et si $\tau>0$ désigne la valeur la plus proche de t' pour laquelle on a de nouveau $\psi=0$, et u_1, z_1 les valeurs cor-

* Les équations des trajectoires de phase γ et γ' sur le plan $\psi=0$, tangentes aux droites $u+(A-1)z\pm B=0$, s'obtiennent facilement à partir de (2).

respondantes de u, z, on a

$$u_0 = u_0(\zeta, \tau) \equiv (A-1)(1+\zeta)\frac{e^{-\tau}-1}{\tau} + \frac{\tau}{2} + A + B - 1;\ z_0 = z_0(\zeta, \tau) \equiv \zeta \qquad (3)$$

$$u_1 = u_1(\zeta, \tau) \equiv (A-1)(1+\zeta)\frac{e^{-\tau}-1}{\tau} - \frac{\tau}{2} + A + B - 1;$$

$$z_1 = z_1(\zeta, \tau) \equiv (1+\zeta)e^{-\tau} - 1$$

Les formules (3) donnent sous une forme paramétrique la transformation du point u_0, z_0 du morceau $G^{(1)}$ en le point u_1, z_1 appartenant à $G^{(2)}$, $\overline{G}^{(2)}$ ou $G^{(3)}$ (transformation S^+)*. Remplaçant dans (3) le paramètre ζ par $-\zeta$ et changeant le signe des coordonnées u et z, nous obtiendrons les formules

$$u_0 = u_0^*(\zeta, \tau),\ z_0 = z_0^*(\zeta, \tau);\ u_1 = u_1^*(\zeta, \tau),\ z_1 = z_1^*(\zeta, \tau) \qquad (3')$$

d'après lesquelles le point u_0, z_0 appartenant à $G^{(3)}$ se transforme en un point u_1, z_1 appartenant à $G^{(2)}$, $\overline{G}^{(2)}$ ou $G^{(1)}$ (transformation S^-).

L'ensemble des transformations S^+, S^-, E et $\overline{E}$ forme la transformation à étudier du plan u, z en lui-même.

4. Supposons que le point $P(u_0, z_0; \zeta, \tau)$ du morceau $G^{(1)}$ se transforme en le point $P_1(u_1, z_1; \zeta, \tau)$ du morceau $G^{(3)}$, et ensuite le point $P_1(u_1, z_1; \zeta', \tau')$ se transforme en le point $u_2 = u_0$, $z_2 = z_0$. On a alors

$$\begin{aligned} u_0(\zeta, \tau) &= u_1^*(\zeta', \tau');\ z_0(\zeta, \tau) = z_1^*(\zeta', \tau') \\ u_1(\zeta, \tau) &= u_0^*(\zeta', \tau');\ z_1(\zeta, \tau) = z_0^*(\zeta', \tau') \end{aligned} \qquad (4)$$

L'étude du système (4) montre qu'il possède seulement des solutions telles que $\tau = \tau' = \tau^* > 0$, $\zeta = -\zeta' = \zeta^* > 0$.

Les valeurs de τ^* et ζ^* sont déterminées par les équations (5):

$$2(A-1)(e^{-\tau}-1) = \tau(1+e^{-\tau})(1-A-B);\ 1-\zeta = (1+\zeta)\,e^{\frac{2(A-1)}{1-A-B}\zeta} \qquad (5)$$

Pour $A+B>1$ ces équations n'ont pas de racine positive, et par conséquent les points invariants considérés n'existent pas. Dans ce qui suit nous distinguons trois cas:

$A+B>1$ (cas I);

$A+B<1$, $[(A+B)(A-B-1)+B]\,e^{\frac{2B}{A+B}} < (A+B)(A+B-1)-B$ (cas II);

$A+B<1$, $[(A+B)(A-B-1)+B]\,e^{\frac{2B}{A+B}} > (A+B)(A+B-1)-B$ (cas III).

5. Dans le cas II chacune des équations (5) possède une racine positive et une seule. L'étude de l'équation caractéristique correspondante montre que P est un col. Cela signifie en particulier que les points qui, lorsque la transformation est réitérée, se rapprochent indéfiniment de P sont situés sur deux courbes L_1 et L_2 (ω-séparatrices du col), formant avec P une courbe analytique en P; les α-séparatrices de P forment également avec P une courbe L. Dans $G^{(3)}$ symétriquement à P, L_1, L_2 et L par rapport à l'origine se trouvent le col $P' \equiv P_1$ et les séparatrices correspondantes.

6. Dans le cas III il existe deux points appartenant à $G^{(2)}$ tels que l'application consécutive des transformations E, S^-, S^+ les transforme en eux-mêmes. Ces point définissent dans l'espace u, z, ψ une trajectoire périodique instable dont font partie les morceaux de trajectoires appartenant à la «lame».

7. Commençons l'étude de la structure globale de l'extension en phase en examinant à l'aide de la transformation de Poincaré la conduite des parties suffisamment éloignées du plan. Introduisons $\eta = \frac{1}{u}$, $\xi = \frac{z}{u}$.

---

* Les paramètres τ et ζ, ceci étant, peuvent prendre n'importe quelle valeur satisfaisant es inégalités

$$\tau > 0,\ \tau^2 \geqslant 2\zeta\,(1-\tau-e^{-\tau})\,(A-1).$$

Des formules pour S^+ et S^- définissent une transformation du plan ξ, η en lui-même; l'axe $\eta=0$, qu'elle transforme en lui-même, correspond à la droite infiniment éloignée du plan u, z (équateur de la sphère de Poincaré). L'étude de cette transformation et de celle analogue, $\eta=\frac{1}{z}$, $\xi=\frac{u}{z}$, montre qu'il existe sur l'équateur de la sphère de Poincaré seulement deux points invariants (aux «bouts» de l'axe u), stables lorsque $A+B>1$ et instables lorsque $A+B\leqslant 1$.

Transportons l'étude sur une des hémisphères de Poincaré. Une partie de l'hémisphère $G^{(1)}$ devient en vertu de la transformation S^+ un certain ensemble $G_1^{(1)}$, dont les points appartiennent à $G^{(2)}$, $\overline{G}^{(2)}$ et $G^{(3)}$. Appliquant aux points $G_1^{(1)}$ les transformations E, E et S^-, on obtient un ensemble de points $\overline{G}_2^{(1)}$, contenant le domaine $G_2^{(1)}$, compris dans $G_1^{(1)}$. L'examen de la conduite des points du domaine $\overline{G}_2^{(1)}$ non contenus dans $G_2^{(1)}$ et de celle des points du

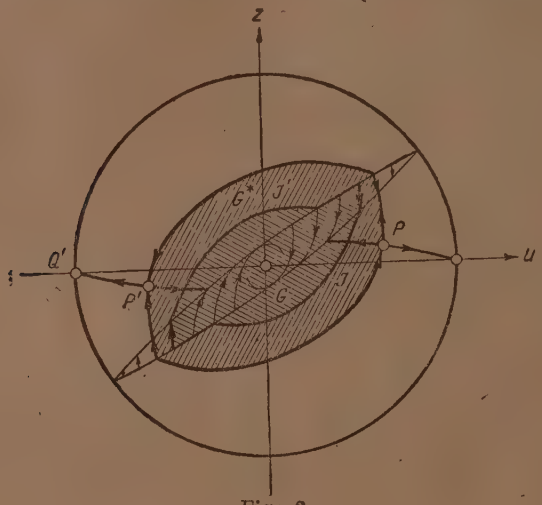

Fig. 2

domaine $G_2^{(1)}$ [effectué à l'aide d'un artifice permettant d'utiliser les raisonnements de Brouwer ([10])] conduit aux conclusions suivantes:

Cas I. La transformation étant réitérée, tous les points du plan tendent vers l'état d'équilibre.

Cas II. Les ω-séparatrices des points invariants P et P' délimitent un ensemble ponctuel G^* tel que si le point initial appartient à G^*, les points transformés convergent vers l'état d'équilibre; si, par contre, le point initial n'appartient ni à G^*, ni à sa frontière, les points transformés s'éloignent à l'infini*. Dans ce cas il n'y a pas sur le plan de points non errants [terminologie de Birkhoff ([11–12])] autres que P, P' et l'origine (fig. 2).

Cas III. Là encore on peut délimiter un domaine G^{**}, qui est un domaine d'attraction pour l'état d'équilibre. Si le point initial n'appartient ni à G^{**}, ni à sa frontière, les points transformés s'éloignent à l'infini. Dans ce cas il existe sur le plan, outre l'origine, d'autres points non errants, appartenant à la trajectoire périodique instable dont il a été question au n° 6.

8. Comme il a été montré plus haut, il convient de distinguer deux cas de convergence du processus de régulation:

I) $A+B>1$ convergence, quelles que soient les conditions initiales;

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* Utilisant certaines propriétés spéciales de la transformation (3), on peut montrer que G^* est un domaine simplement connexe.

2) $A+B<1$ convergence, si les conditions initiales se trouvent respectivement dans G^* (cas II) ou dans G^{**} (cas III).

Si les conditions initiales n'appartiennent pas à ces domaines ou à leurs frontières, le processus de régulation diverge, le dispositif automatique n'atteint pas son but*.

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* Appellons domaine G la partie de G^* (ou G^{**}) dont les points, dès la première transition, entrent dans le domaine $\overline{G}^{(2)}$ et tendent ensuite vers l'état d'équilibre sans quitter la «lame». Les frontières du domaine G s'obtiennent facilement sous forme paramétrique. La connaissance des frontières du domaine G permet une évaluation grossière des frontières du domaine G^* (ou G^{**}). Le calcul exact des frontières des régions de convergence présente certaines difficultés (surtout dans le cas II).

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PHYSICS

ON DETERMINATION OF SYMMETRY OF NUCLEAR SPIN-FUNCTIONS

By I. GODNEV

(Communicated by V. A. Fock, Member of the Academy, 20.I.1944)

For determining the weights of rotational energy levels in polyatomic molecules and solving a number of other problems it is necessary to know the symmetry of the nuclear spin-functions of the molecule.

Wilson (1) has turned to the theory of groups for a method of calculation of rotational levels and shown that the problem is considerably simplified by establishing the symmetry of wave functions with respect to the group $\mathfrak{R}$ of such permutations of identical nuclei as may be effected by the rotation of the molecule. In determining the symmetry of nuclear spin-functions Wilson (1, 2) used a method based on a direct computation of the traces of matrices of the representation of the group $\mathfrak{R}$. In this paper I wish to point out that there is another method by which the same purpose can in certain cases be attained in a less laborious way.

Let $\mathfrak{S}_n$ be the group of all permutations n of similar nuclei, having a spin (symmetric group), and $\mathfrak{R}$, as indicated above, the group of the permutations effected by rotations that bring the molecule in a position identical to its initial position. With respect to $\mathfrak{S}_n$ $\mathfrak{R}$ is a subgroup, its order is equal to the symmetry number of the molecule σ.

If the nuclear spin can take α values, there should exist α^n independent nuclear spin-functions forming the basis of an α^n-dimensional representation D of the groups $\mathfrak{S}_n$ and $\mathfrak{R}$. The symmetry of nuclear spin-functions is determined by the irreducible representations contained in D.

On the method of Wilson we have to compute the traces of the matrices of the representation D and fulfil its reduction to the irreducible representations of the group $\mathfrak{R}$. But to carry out this reduction we need not compute the traces of the matrices of D. In fact, the resolution of D into irreducible representations of a symmetrical group is possible in the general form

$$D = b_0 B_0 + b_1 B_1 + \ldots + b_s B_s \qquad (1)$$

where the coefficients b_k can be computed by means of explicit formulae.

Generally speaking, the representations $B_0, B_1, \ldots, B_s$ are not irreducible with respect to the group $\mathfrak{R}$, which is a subgroup of $\mathfrak{S}_n$. The reduction of these representations, which is easily made by using tables of their characters, gives

$$B_k = g_{k1}\Gamma_1 + g_{k2}\Gamma_2 + \ldots + g_{kp}\Gamma_p \qquad (2)$$

Substituting in (1) the values of B_k, we shall find the sought-for symmetry of nuclear spin-functions

$$D = a_1\Gamma_1 + a_2\Gamma_2 + \ldots + a_m\Gamma_m \qquad (3)$$

Thus, is case $\alpha = 2$, *e. g.* for molecules containing n atoms of hygrogen, the representations B_k contained in D correspond (3) *partitio numerorum* to $(n-k)+k$ for integral values of k ranging from zero to $k \leqslant {}^1/_2\, n$. The number of such representations for n even and odd is $\frac{n}{2}+1$ and $\frac{n+1}{2}$, respectively. The dimensionality of B_k is $\binom{n}{k} - \binom{n}{k-1}$ and its multiplicity is $b_k = n + 1 - 2k$. Carrying out the reduction of these representations, whose characters are known, it is easy to find the coefficients g_{ki} in formula (2), whereby the a_k will be also established.

In the case of benzene, for instance, $n = 6$; $\sigma = 12$; the irreducible representations of $\mathfrak{R}$ are(4) A_1, A_2, B_1', B_2', E^*, E_*^*.

In virtue of (1)

$$D = 7B_0 + 5B_1 + 3B_2 + B_3 \qquad (4)$$

On the reduction of the representations B_0, B_1, B_2, B_3, whose characters are to be found, for instance, in a paper by Frobenius (5) where these representations are denoted by $\chi^{(0)}$, $\chi^{(2)}$, $\chi^{(4)}$ and $\chi^{(5)}$, respectively, we have

$$\begin{aligned} B_0 &= A_1 \\ B_1 &= B_1' + E^* + E_*^* \\ B_2 &= 2A_1 + B_2' + 2E^* - E_*^* \\ B_3 &= A_2 + 2B_1' + E_*^* \end{aligned} \qquad (5)$$

From (4) and (5) follows

$$D = 13A_1 + A_2 + 7B_1' + 3B_2' + 11E^* + 9E_*^* \qquad (6)$$

and this result coincides with that obtained by the computation of traces of D and direct reduction (1).

For ethane $n = 6$, $\sigma = 18$, the characters of the irreducible representations Γ_l of the group $\mathfrak{R}$ are given by Wilson (2).

The reduction of B_k gives

$$\begin{aligned} B_0 &= \Gamma_1 \\ B_1 &= \Gamma_2 + \Gamma_7 + \Gamma_8 \\ B_2 &= \Gamma_1 + \Gamma_3 + \Gamma_4 + \Gamma_7 + \Gamma_8 + \Gamma_9 \\ B_3 &= \Gamma_2 + \Gamma_5 + \Gamma_6 + \Gamma_9 \end{aligned} \qquad (7)$$

Substituting in (4) the values of B_k we have

$$D = 10\Gamma_1 + 6\Gamma_2 + 3\Gamma_3 + 3\Gamma_4 + \Gamma_5 + \Gamma_6 + 8\Gamma_7 + 8\Gamma_8 + 4\Gamma_9. \qquad (8)$$

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PHYSIQUE

PERMÉABILITÉ MAGNÉTIQUE TRANSVERSALE

Par G. GORELIK

(Présenté par L. I. Mandelstam, de l'Académie, le 1.X.1943)

Etudiée pour la première fois par R. Gans (1), la perméabilité magnétique transversale a acquis un intérêt particulier depuis les recherches de Harrison, Turney, Rowe et Gollop (2) et de Webb (3) sur l'impédance de fils ferromagnétiques aimantés longitudinalement, parcourus par des courants alternatifs. La variation de la perméabilité transversale en fonction du champ longitudinal est utilisée dans le «magnétomètre à impédance» de Harrison et Rowe (4). La variation de la perméabilité transversale en fonction des champs longitudinal et transversal est susceptible d'applications diverses en radioélectricité. A celles énumérées par Webb (3) nous pouvons ajouter: réalisation de systèmes à paramètres variant périodiquement, modulation en fréquence, changement de fréquence.

Cette Note contient quelques considérations théoriques (très élémentaires) sur la variation de la perméabilité transversale en fonction des champs longitudinal et transversal, et leur comparaison avec les résultats expérimentaux obtenus par l'auteur en collaboration avec I. Borstein, V. Kouteïnikov, A. Lubina, X. Goronina et I. Joukova.

§ 1. Considérations théoriques

1. Lorsque le champ longitudinal est faible, la relation entre l'induction et le champ transversaux est à peu près la même qu'entre l'induction et le champ longitudinaux. Il y a donc, dans la direction transversale: hystéresis et non-linéarité fortement marqués.

2. En cas de champ longitudinal très intense, lorsqu'en absence de champ transversal il y a saturation longitudinale, la perméabilité transversale peut être calculée très simplement. En effet, les énergies cristalline et magnétoélastique sont alors négligeables par rapport à l'énergie du champ extérieur, et l'aimantation spontanée, ainsi que l'induction de saturation, sont parallèles au champ (le corps est supposé isotrope). On a donc

$$\frac{B_\perp}{H_\perp} = \frac{B_\parallel}{H_\parallel}$$

$B_\perp$, $H_\perp$ étant l'induction et le champ transversaux, $B_\parallel$, $H_\parallel$ l'induction et le champ longitudinaux.

Eliminant $B_\parallel$ à l'aide de la relation

$$B_\parallel = \sqrt{B_s^2 - B_\perp^2}$$

il vient

$$B_\perp = B_s \left\{ \frac{H_\perp}{H_\parallel} - \frac{1}{2}\left(\frac{H_\perp}{H_\parallel}\right)^3 + \ldots \right\}$$

et la perméabilité transversale

$$\mu_\perp \equiv \frac{B_\perp}{H_\perp} = \frac{B_s}{H_\parallel}\left\{1 - \frac{1}{2}\left(\frac{H_\perp}{H_\parallel}\right)^2 + \ldots\right\}.$$

La relation entre l'induction et le champ transversaux est donc ici univoque et d'autant plus linéaire que le champ longitudinal est intense. Pour

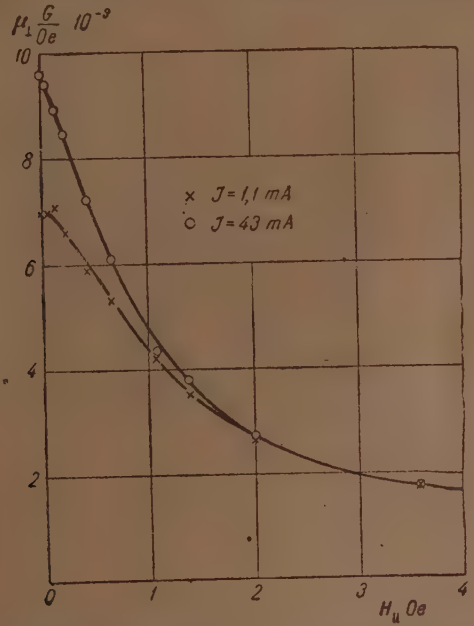

Fig. 1

une valeur fixée du champ longitudinal la perméabilité transversale décroît lorsque le champ transversal augmente.

§ 2. Comparaison avec l'expérience

Les mesures ont été faites sur un alliage de haute perméabilité du nickel et du fer avec l'addition de cuivre et de molybdène. La perméabilité transversale était déterminée à partir de l'impédance Z, mesurée à la fréquence 16 000, de fils (longueur: environ 40 cm, diamètre: 0.4 mm) placés dans un champ statique longitudinal, à l'aide de la formule bien connue

$$\frac{Z}{R_0} = \frac{ka}{2} \cdot \frac{J_0(ka)}{J_1(ka)}$$

R_0 étant la résistance à la fréquence 0; a, le rayon du fil; $k = (1+j)\,2\pi\sqrt{\mu_\perp \nu \sigma_m}$; ν, la fréquence; σ_m, la conductibilité en unités CGSM; J_0, J_1, les fonctions de Bessel d'ordre 0 et 1 (en cas de relation non-linéaire entre $B_\perp$ et $H_\perp$ cette formule définit une perméabilité transversale effective).

La fig. 1 représente en fonction du champ longitudinal la perméabilité transversale mesurée pour deux intensités efficaces du courant dans le fil: 1.1 mA (courbe inférieure) et 43 mA, ce qui correspond à la surface du fil à des champs transversaux d'amplitudes 0.015 et 0.6 Oe environ.

La fig. 2 représente en fonction du champ longitudinal l'inverse de la perméabilité transversale. Les signes différents marquant les points expérimentaux correspondent à des fils différents de même composition, ayant subi le même traitement thermique. L'examen de ces figures montre que:

1) Le champ longitudinal augmentant, la variation de $\mu_\perp$ en fonction de l'ntensité du courant devient de moins en moins marquée et disparaît, à la précision de nos mesures, aux environs de $H_\parallel = 2.5$ Oe;

2) dans un champ longitudinal intense $1/\mu_\perp$ est proportionnel à $H_\parallel$, le quotient de ces grandeurs étant $\frac{0.002}{15} = \frac{1}{7500}$ Gauss^{-1}.

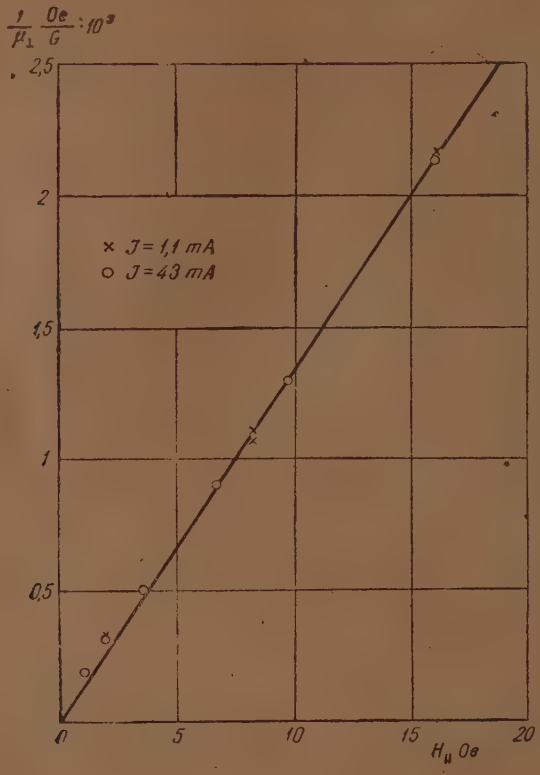

Fig. 2

La mesure de l'induction de saturation par la méthode balistique usuelle donne, à l'erreur d'observation près, la valeur $B_s = 7500$ Gauss.

Nous pouvons conclure que l'expérience confirme nos considérations théoriques. (L'augmentation de la perméabilité transversale effective, accompagnant celle du courant qu'indique la fig. 1 pour les petits valeurs de $H_\parallel$, n'a rien de commun avec la variation de $\mu_\perp$ en fonction de $H_\perp$ indiquée au § 1, n° 2. Cette augmentation est due, probablement, en majeure partie, à celle des pertes d'hystéresis).

Soumis à l'action d'un champ longitudinal de l'ordre de quelques Oersteds nos fils réunissent donc, dans la direction transversale, les propriétés suivantes: 1) perméabilité de l'ordre de $10^3 \frac{\text{Gauss}}{\text{Oersted}}$; 2) haut degré de linéarité; 3) possibilité de modifier la perméabilité dans de grandes limites en faisant varier le champ longitudinal. Cet ensemble de propriétés présente un intérêt évident pour les applications indiquées plus haut.

Je suis très reconnaissant à N. D. Papalexi, membre de l'Académie, et au docteur V. L. Ginsburg pour nos entretiens concernant le sujet de cette Note.

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Gorki.

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GEOPHYSICS

ON A RAPID METHOD OF ATMOSPHERIC TRANSPARENCY DETERMINATION

By E. PIASKOWSKAIA-FESSENKOWA

(Communicated by B. G. Fessenkow, Member of the Academy, 6.XII.1943)

The distribution of brightness over a diurnal sky may be represented by the following simple formula, given by B. Fessenkow (1), where an arbitrary indicatrix of (ϑ) characterizes the whole atmosphere, the Earth's surface is assumed to be flat and the scattering is of the first order:

$$I = L\frac{\sigma}{k} f(\vartheta)\frac{\sec z}{\sec z - \sec z_\odot}\left(p^{\sec z_\odot} - p^{\sec z}\right) = L\frac{\sigma}{k} f(\vartheta)\varphi(\sec z_\odot, \sec z) \qquad (1)$$

ϑ is the angle between the incident and diffused rays; L, the illumination from the Sun outside of the atmosphere; z, $z_\odot$, the zenithal distances of the observed point of the sky and the Sun; p, the coefficient of atmospheric transparency; k, the absorption coefficient.

Consequently, the sky brightness in a given direction is always a product of two functions depending on ϑ and p, z, respectively. This is incorrect if the atmosphere at different levels is characterized by different indicatrices. In the case of two indicatrices $f_1(\vartheta)$, $f_2(\vartheta)$ we have

$$I = L\frac{\sigma}{k} f_1(\vartheta)\frac{\sec z}{\sec z - \sec z_\odot}\left(p^{\sec z_\odot} - p^{\sec z}\right)(1+\Delta) \qquad (2)$$

where Δ is a certain function of f_1, f_2. Nevertheless it is possible so to choose the value of ϑ that $\Delta = 0$ and the brightness distribution at a corresponding angular distance from the Sun will be the same as in the case of a single indicatrix.

Numerous photometrical determinations of the sky brightness were made in 1942 in the mountain region near Alma-Ata at a height of 1800 meters by an expedition of the Institute of Astronomy and Physics.

I have found these observations be strictly consistent with the assumption of a single $f(\vartheta)$. Consequently, the formula (1) may be used, and the parameter p must be considered as a real coefficient of atmospheric transparency.

This parameter may be easily determined if the values of brightness I_1, I_2 for two points of the sky lying at an equal angular distance ϑ are known. It is sufficient to consider the ratio

$$\frac{I_1}{I_2} = \frac{\left(1 - p^{m_1 - m_\odot}\right)(m_2 - m_\odot)\, m_1}{\left(1 - p^{m_2 - m_\odot}\right)(m_1 - m_\odot)\, m_2} \qquad (3)$$

if $m_1 > m_\odot$, $m_2 > m_\odot$; or

$$\frac{I_1}{I_2} = \frac{p^{m_\odot - m_2}(1-p^{m_1-m_\odot})(m_\odot - m_2)\, m_1}{(1-p^{m_\odot - m_2})(m_1 - m_\odot)\, m_2} \tag{4}$$

if $m_1 > m_\odot$; $m_2 < m_\odot$; or at last

$$\frac{I_1}{I_2} = \frac{p^{m_1 - m_2}(1-p^{m_\odot - m_1})(m_\odot - m_2)\, m_1}{(1-p^{m_\odot - m_2})(m_\odot - m_1)\, m_2} \tag{5}$$

if $m_1 < m_\odot$, $m_2 < m_\odot$.

In all of these cases it is assumed that $m_1 > m_2$, m being the so-called atmospheric mass which is very nearly the same as $\sec z$. Let us take

$$p = 1 - p'.$$

Then from (3)

$$\left(\frac{I_1}{I_2} m_2 - m_1\right) - \frac{1}{2!}\left[\frac{I_1}{I_2} m_2 (m_2 - m_\odot - 1) - m_1 (m_1 - m_\odot - 1)\right] p' +$$
$$+ \frac{1}{3!}\left[\frac{I_1}{I_2} m_2 (m_2 - m_\odot - 1)(m_2 - m_\odot - 2) - m_1 (m_1 - m_\odot - 1)(m_1 - m_\odot - 2)\right] p'^2 -$$
$$- \ldots + (-1)^{n-1} \frac{1}{n!}\left[\frac{I_1}{I_2} m_2 (m_2 - m_\odot - 1) \ldots (m_2 - m_\odot - n + 1) -\right.$$
$$\left. - m_1 (m_1 - m_\odot - 1) \ldots (m_1 - m_\odot - n + 1)\right] p'^{n+1} + \ldots = 0 \tag{6}$$

This series is convergent rapidly; its first three terms are sufficient for exact determination of p'. In analogous manner one finds from (4) that

$$\left(\frac{I_1}{I_2} m_2 - m_1\right) - \left\{\frac{1}{2!}\frac{I_1}{I_2} m_2 (m_\odot - m_2 - 1) - m_1\left[\frac{1}{2!}(m_1 - m_\odot - 1) + (m_\odot - m_2)\right]\right\} p' +$$
$$+ \left\{\frac{1}{3!}\frac{I_1}{I_2} m_2 (m_\odot - m_2 - 1)(m_\odot - m_2 - 2) - m_1\left[\frac{1}{3!}(m_1 - m_\odot - 1)(m_1 - m_\odot - 2) +\right.\right.$$
$$\left.\left. + \frac{1}{2!}(m_\odot - m_2)\,|\,(m_1 - m_\odot - 1) + (m_\odot - m_2 - 1)\,|\right]\right\} p'^2 - \ldots = 0 \tag{7}$$

The formulae (6) and (7) are applicable to the points of the sky lying at the same angular distance from the Sun. The calculation is very simple if $m_1 - m_\odot$, $m_2 - m_\odot$ are whole numbers.

By way of example let us consider the observations made on September 3; 1942 of the diurnal sky brightness.

Choosing $m_1 - m_\odot = 2$; $m_\odot - m_2 = 1$; $m_\odot = 3.816$, and making use of the formula

$$p^2 + p - 2\frac{I_1}{I_2}\frac{m_2}{m_1} = 0$$

one obtains after substitution of the corresponding values for I_1 and I_2

$$p = 0.740 \pm 0.006.$$

On the other hand, putting $m_1 - m_\odot = 1$, $m_\odot - m_2 = 2$, one finds from the equation

$$2p^2 - \frac{I_1 m_2}{I_2 m_1} p - \frac{I_1 m_2}{I_2 m_1} = 0$$

that

$$p = 0.731 \pm 0.005.$$

Moreover, if $m_1 - m_\odot = m_\odot - m_2 = q$, where q is a positive number, we have

$$p^q = \frac{I_1}{I_2}\frac{m_2}{m_1}$$

Putting, for instance, $q=2$, one obtains for the same days of observations

$$p = 0.733 \pm 0.004.$$

If again $q=1$, then

$$p = \frac{I_1}{I_2}\frac{m_2}{m_1}$$

whence

$$p = 0.741 \pm 0.009.$$

It may be seen that the different values for the coefficient of transparency, derived by considering the brightness at different pairs of points of the diurnal sky, are equal within the mean error of observation. By this procedure the coefficient of trapnsarency is determined with high rapidity, if there is a photometer to compare the brightness of the sky simultaneously at two points located at the same angular distance from the Sun. The result under consideration may simply be read from the instrument.

The calculation is extremely simple if one point of the sky is taken at zenith ($m_2 = 1$) and the other below the Sun in such a manner that

$$m_1 = 2m_\odot - 1.$$

Then

$$p^{m_\odot - 1} = \frac{1}{2m_\odot - 1}\frac{I_1}{I_2}$$

At $m_\odot = 2$, one obtains a simple expression, namely

$$p = \frac{1}{3}\frac{I_1}{I_2}$$

The described method gives a reliable value of p. Indeed, the relative distribution of brightness over the diurnal sky is practically independent of the light scattering of higher orders (2) or of the reflexion of light from the ground.

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PHYSICAL CHEMISTRY

EFFECT OF ADSORBED AMMONIA ON TRIBOELECTRICITY POTENTIAL OF COMPRESSED GASES

By M. S. FURMAN, I. P. SIDOROV and J. S. KAZARNOVSKY

(Communicated by A. N. Frumkin, Member of the Academy, 29.XI.1943)

The authors have observed charges of triboelectricity to arise when gases compressed to about 500—600 atm. are let out through a narrow nozzle, 1 mm² in section. In potential these charges amount to values sufficient to cause puncture. Here is danger if one has to deal with gaseous mixtures possessing explosive properties.

We thought it expedient to explore the possibilities for the elimination of this danger. The most efficient way was to create such conditions as would increase the breakdown power or accelerate the flow of charges beyond the rate of their accumulation.

The puncture voltage of gases is increased by the introduction of CCl_4, benzene, hexafluoride of sulphur, freon and other substances. Yet in our case the results required could not possibly be attained in that way, for the substance added condensed under pressure, and the presence of liquid drops sharply increases the value of the potential arising when a compressed gas is let out.

The other way is to increase the rate of flow of the charges. Here the properties of the water vapour might be turned to good account, but its concentration under so high a pressure would be too insignificant. So we decided to use ammonia for that purpose. At first we found that ammonia produced a sharp rise of the potential which again was due to its condensation under high pressure. But the results were different when the vessel had been treated with ammonia.

We found that after the adsorption of ammonia by the surface of the vessel, a static voltmeter of the Hartman-Braun type gave no response whatever to the letting out of the compressed gas through a narrow nozzle.

Without ammonia, other conditions being equal, the potential reached the following values:

Pressure, kg/cm²	400	300	200
Potential, Volt	1 100	510	40

Relative humidity of the atmosphere, 22 per cent
Gas tested, nitrogen

If the surface of the vessel is first treated with H_2S and then with NH_3, no disappearance of potential is observed. This is probably because the surface is poisoned. So there is definite evidence of a complete reversibility of the effect caused by the application of ammonia to the surface. We do not venture an explanation of the mechanism of this phenomenon at this juncture.

Institute of Nitrogen Industry.

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CHEMISTRY

STEREOCHEMISTRY OF ANHYDRIDES

By J. V. KHODAKOV

(Communicated by V. G. Chlopin, Member of the Academy, 28.VIII.1943)

In accordance with the subordination scheme

crystallochemistry — stereochemistry

composition — structure — chemical properties

structure in stereochemistry is no longer the aim of investigation, as in crystallochemical studies, but rather a means for the elucidation of that complex of properties of the substance which constitutes its chemical individuality. Assuming, contrary to the opinion that has been prevalent ever since the times of Werner, that the coordination number that appears in a metalloid in the case of compounds of the second order (acids and salts) is also retained by its anhydride, we take for our starting point two statements established on the basis of the material available in the crystallochemistry of silicates (1).

1) Every high valency atom with the coordination number 4 tends to form around itself a tetrahedron composed of oxygen ions.

2) These tetrahedra avoid having common edges, not to speak of faces (the so-called third principle of Pauling).

The problem we have set ourselves is to find such configurations for the anhydride X_nO_m (X = Cl, S, P, Si) as would satisfy these conditions, and after comparing the conclusions with the experiment to apply them for the interpretation of the paradoxical facts so often encountered in the complicated chemistry of sulphuric and phosphoric anhydrides, which substances are of fundamental importance for smoke screening.

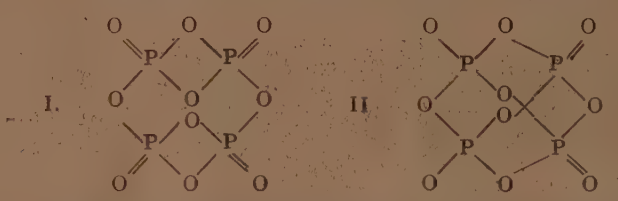
In order that the conditions of valence and coordination be fulfilled simultaneously, it is necessary that all (Si) or part (3 in P_2O_5, 2 in SO_3, and 1 in Cl_2O_7) of the four oxygen ions grouping around any cation X^1 should enter simultaneously in the composition of the tetrahedra formed around the neighbouring cations X^2, X^3... In the case of anhydrous perchloric acid this leads to a single solution, the closed molecule of $O_2ClOClO_3$. Therefore for Cl_2O_7 only one modification is known, not polymerizing and volatile (2).

In the case of sulphuric anhydride one obtains instead of a molecule a link with not saturated valencies $-SO_2O-$. Such links may combine to form either a ring $(SO_3)_n$ ($n \geqslant 3$), or an infinitely extended chain, as is the case with cellulose, silk, asbestos and others. This may be taken to mean that sulphuric anhydride is capable of polymerization and in its polymerized form presents a prototype of fibrous substances. Hence the dependence of the properties of SO_3 («chemical chameleon») (3) upon the previous history of the preparation and its transformation in the course of time from an ice-like form α-SO_3, easily fusible and volatile,

into an asbestos-like β-SO_3, white threads woven into a felt-like texture having a silky lustre; they are not easily cut with a knife and assume the shape of wet cotton-wool previous to thermal destruction (4).

The mechanism of polymerization of SO_3 may be of two different kinds. The ring $(SO_3)_n$ may widen, thereby involving new links, but cannot possibly break, for in this case the terminal atoms S and O would prove to have unsaturated valencies. The transformation of a ring-like structure into a linear structure is not possible without participation of a foreign substance, *e. g.* water which takes upon itself the task of saturating the valencies of the disjoining atoms. Then a molecule of the acid $S_{n-2}O_{3n-1}(SO_2OH)_2$ is obtained instead of the polymolecule $(SO_3)_n$. But if n is large, the deviation from the empirical formula for SO_3 will not be considerable. If we suppose that the water is involved in the process of polymerization from the very beginning, we arrive at a «catalytic polymerization», whose initial point is the molecule H_2SO_4, and each of the successive acts is reduced to a replacement of H by a sulpho-group ($H_2SO_4 \rightarrow H_2S_2O_7 \rightarrow H_2S_3O_{10}$ ...). The influence of traces of water on the transformation α-$SO_3 \rightarrow$ β-SO_3 is recognized as an established fact (3). Without them no transformation takes place even after many a year has elapsed (4). That the elements of water enter into the composition of β-SO_3 in the ratio of $H_2O : SO_3 \cong 1 : 1000$, has also been proved by analysing its samples (6, 7).

In the modifications of phosphoric anhydride three ions of oxygen out of a number of four, must each be divided between two different atoms of phosphorus. This is possible in three different ways. Either all of the three oxygen ions are bound by a P^1 ion with the same second ion P^2, or two of them with one P ion, and the third with another, or at last all the three O'' ions are bound with three different P ions (P^2, P^3 and P^4). The first solution leads to a closed molecule OPO_2PO, while the remaining two yield the links $OP^1(O_2P^2)(OP^3)$ and $OP^1(OP^2)(OP^3)(OP^4)$. Both solutions (if we take into consideration that the number of P atoms cannot possibly be odd) imply the presence in the molecule of at least four P atoms, which leads to the isomeric structures



Of them II, contrary to I, adopted in chemistry, is the simplest possible structure of phosphoric anhydride, satisfying besides the coordination conditions, the third principle of Pauling. In space it presents a tetrahedron formed by four PO_4 tetrahedra slightly flattened toward the centre of the molecule. Each of these tetrahedra has three of its vertices joined to the vertices of the remaining three. The oxygen ions are arranged into the closest possible packing (cubic packing), so that Bragg's principle is also fulfilled (1). There is every reason therefore to expect that in the form polymerized to the least extent, *e. g.* in vapour, the sulphuric anhydride should be represented by molecules of just such a structure. In fact, its vapour up to very high temperatures corresponds to the formula P_4O_{10} (8), and its electronographic investigation by Hampson and Stosick (9), made recently, has confirmed the structure derived by me.

Comparing the structures of P_4O_6 and P_4O_{10} (see figure *) one is struck by the

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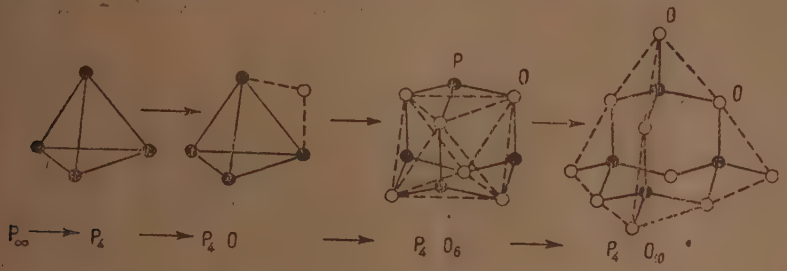
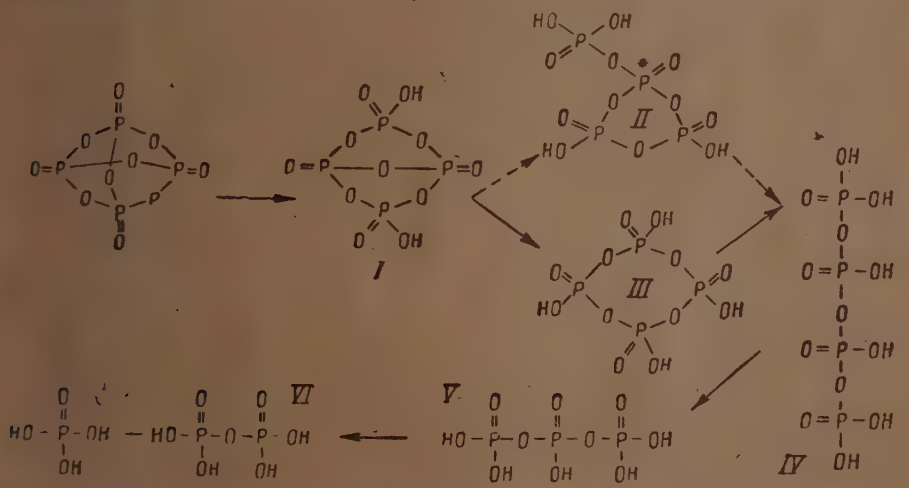
* In P_4O_6: P — O = 1.65 Å, ∠OPO = 99°; in P_4O_{10}: P — O = 1.62 Å, P — O* = 1.39 Å, ∠OPO = 101.5°, ∠OPO* = 116.5°, whereas the structure of P_4O_{10}, calculated by me as a clue to the stereochemistry of oxygen compounds of phosphorus before I became acquainted with Hampson and Stosick's paper, implies that P — O = 1.56 Å (as in KH_2PO_4); P—O* = 1.4 Å (< P — O on account of the correction for coordination) and ∠OPO = ∠OPO* = 109.5°.

close correspondence between them. The former is converted into the latter without any rearrangement whatsoever, even with the atoms scarcely being displaced, but merely because of the fact that four additional oxygen atoms (O*) take the position pre-assigned, as it were, to them near the phosphorous atoms. Hence the easy oxidability of P_4O_6 into P_4O_{10}.



From a stereochemical point of view the hydration of an anhydride should consist in a successive disjunction of the «welded» vertices of XO_4 tetrahedra. In the case of Cl_2O_7 this leads directly to $HClO_4$. In the case of β-SO_3 the hydration should be reduced to disintegration of longer molecules into shorter ones up to H_2SO_4. And in the case of P_2O_5, if one proceeds from the molecule of P_4O_{10} in the vapour state, the successive stages of hydration may be expressed by the following scheme:



«Swelling» stage: I, hypothetic primary acid; II and III, meta-acids; IV, acid of Rakusin. «Splitting» stage: V, acid of Schwarz; VI, pyroacid

It may be seen from the scheme that the hydration of P_2O_5 is composed of a swelling stage and a splitting stage, and may proceed along two different ways, of which the one, leading through meta-acid II, is scarcely probable, in view of the presence in it of anhydrous phosphorus P*+. It is also obvious that a number of intermediate forms, the three known polyacids inclusive, should arise in the process of hydration. Still more complicated must be the run of hydration of polymerized forms of P_2O_5, represented by infinitely extended lattices of «welded»

* At the same time one may conclude from the thermal dehydration of H_3PO_4 which does not proceed beyond the formation of meta-acid, that every P atom tends to retain at least one hydroxyl group.

PO_4 tetrahedra. Here the swelling stage should include typical colloid states, gels and sols, as is the case with the cynematographically inverted processes of polymerization in silicic acids.

The experimental data on the hydration of P_2O_5 are incomplete, obscure and conflicting. What has been confirmed by the pre-war work ([10],[11]) is the appearance of $H_4P_2O_7$, which point had been a matter of dispute ever since the times of Bercelius; then the presence of a preceding colloidal stage in the case of the non-volatile form of P_2O_5 and of the splitting stage in the case of the volatile form, including some acids, unknown for the time being([10]); and, finally, the fact that the variation of electroconductivity with time proceeds differently with the meta-acids of the crystalline and of the «amorphous» P_2O_5([11]).

Thus, it is in the involved chemistry of anhydrides which for more than a hundred years has defied all attempts to unravel it by empirical investigation not guided or coordinated by theoretical ideas, that Wollaston's words about the inadequacy of arithmetical relations for explaining the interaction between the atoms, and the need of a three-dimensional geometric representation ([12]) have justified themselves.

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GEOCHEMISTRY

ON THE PROBLEM OF COBALT, NICKEL AND COPPER CONTENT OF SOILS

By D. P. MALIUGA

(Communicated by W. I. Vernadsky, Member of the Academy, 25.VII.1943)

J. Forchhammer (1) was the first to discover in 1855 cobalt, nickel and copper in some soils of Denmark. After this but sparse information appeared concerning the cobalt, nickel and copper content in soils, and it was not unless the publication of the well-known work by W. I. Vernadsky (2) on the wide occurrence of nickel in plants, that appeared works by Bertrand (3), McHargue, Kinson, Hibbard, Slater, Porfiriev (4), Vinogradov (5), and of several others, on the quantitative determination of cobalt, nickel and copper in the soils of the USSR, America, France and other countries. Data on cobalt, nickel and copper content of soils were, however, rather fragmentary ones and mostly connected with the study of the influence of these elements upon the development of plants, with the study of endemics, *etc.*

The object of the present study was to analyse the distribution of cobalt, nickel and copper in soils from a geochemical standpoint. The study was subdivided into a number of points: (a) the study of the distribution of cobalt, nickel and copper in all the chief soil zones of the USSR; (b) determination of the average content (in clarkes, after A. E. Fersman) of cobalt, nickel and copper in soils connected with the study of biogeochemical provinces rich or poor in cobalt, nickel or copper; (c) the study of soils in connexion with the areas of dissemination of cobalt and nickel deposits.

The study involved materials classed according to latitude (collection of soils made by the Biogeochemical Laboratory of the Academy of Sciences of the USSR in 1938). A determination of cobalt, nickel and copper was made in the swamp, podzol, steppe [chernozem, chestnut, serozem (grey earth) soils of semi-arid regions], saline soils, krasnozem (red earth) soils of subtropical areas and in other soils of the Soviet Union.

Analysis of the Soils. Two or three grams of a finely ground soil acidulated with sulphuric acid were ignited in a quartz dish in an electric muffle at 450° and then dissolved for a long time in aqua regia until the areno-argillaceous residue was completely discoloured. If coloured, sparingly soluble minerals were still present, the residue was fused with sodium carbonate and tested for the content of metals with rubeanic acid. The solution obtained, upon the usual separation of silicic acid and addition of 1—2 g of citric acid, was neutralized with a small excess of ammonia, and 2—3 ml of a 0.5 per cent alcohol solution of rubeanic acid (6) ($NH_2CS \cdot SCNH_2$, diamide-dithio-oxalic acid) were added to the solution. After several hours' standing, the precipitated rubeanates of cobalt, nickel and copper (zinc and cadmium) were filtered off and dissolved with hot 20 per cent nitric acid. Cobalt was separated with α-nitroso-β-naphtol after Ilinsky and Knorre(7), and the solution was tested for its nickel and copper content by the polarographic method. Cobalt was determined separately by the same method.

Data on the cobalt, nickel and copper content of soils are tabulated below.

Soils	Place of sampling	Content in % to dry soil			Ratio
		Co	Ni	Cu	Co : Ni : Cu
Peat-gley soil	Swampy tundra, Chibiny region				
0—15 cm, hor. A_0		4.9×10^{-4}	2.6×10^{-3}	2.3×10^{-3}	1 : 5 : 4.6
15—25 cm » A_1		1.5×10^{-4}	1.0×10^{-3}	1.1×10^{-3}	1 : 6. 7 : 7.3
25—50 cm » A_2		2.2×10^{-4}	1.2×10^{-3}	1.3×10^{-3}	1 : 5.5 : 6
Medium-podzol-loamy light soil	Moscow region, forest of the Timiriazev Agr. Acad.				
0—18 cm, hor. A_1		1.4×10^{-3}	2.4×10^{-3}	8.0×10^{-4}	1 : 1.7 : 0.6
18—34 cm » A_2		9.3×10^{-4}	2.2×10^{-3}	5.5×10^{-4}	1 : 2.4 : 0.6
34—60 cm » B_1		6.6×10^{-4}	1.6×10^{-3}	5.4×10^{-4}	1 : 2.4 : 0.8
60—110 cm » B_2		8.3×10^{-4}	2.2×10^{-3}	6.8×10^{-4}	1 : 2.7 : 0.8
Leached grey forest soils	Tula region, Yasnaya Polyana				
0—5 cm, hor. A_1		8.1×10^{-4}	2.7×10^{-3}	1.2×10^{-3}	1 : 3.3 ; 1.5
20—25 cm » A_2		6.4×10^{-4}	1.7×10^{-3}	1.0×10^{-3}	1 : 2.7 : 1.6
40—45 cm » B_1		6.7×10^{-4}	2.2×10^{-3}	9.4×10^{-4}	1 : 3.3 : 1.4
75—80 cm » B_2		7.8×10^{-4}	2.2×10^{-3}	8.4×10^{-4}	1 : 2.8 : 1.1
100—105 cm » C		7.2×10^{-4}	2.1×10^{-3}	7.6×10^{-4}	1 : 2.8 : 0.9
Chernozem clayey, common	Talovaya station, Voronezh region				
0—5 cm, hor. A_1		6.7×10^{-4}	3.9×10^{-3}	1.7×10^{-3}	1 : 6 : 2.5
24—32 cm » A_2		5.6×10^{-4}	3.3×10^{-3}	1.2×10^{-3}	1 : 6 : 2.
80—88 cm » B		7.5×10^{-4}	3.5×10^{-3}	1.9×10^{-3}	1 : 5 : 2.5
128—144 cm » C		1.2×10^{-3}	4.8×10^{-3}	1.6×10^{-3}	1 : 4 : 1.4
Chernozem clayey, Near-Azov Sea type	Salsk station, Rostov region				
0—5 cm, hor. A_1		1.3×10^{-3}	7.7×10^{-3}	2.3×10^{-3}	1 : 6 : 1.8
40—45 cm » A_2		1.0×10^{-3}	4.6×10^{-3}	1.3×10^{-3}	1 : 4.6 : 1.3
90—95 cm » B		1.1×10^{-3}	6.3×10^{-3}	1.7×10^{-3}	1 : 5.7 : 1.6
130—135 cm » C		1.0×10^{-3}	4.5×10^{-3}	1.7×10^{-3}	1 : 4.5 : 1.7
Light-chestnut soil	Semichnaya station, Rostov region				
0—5 cm, hor. A_1		1.2×10^{-3}	5.9×10^{-3}	1.4×10^{-3}	1 : 5 : 1.2
30—35 cm » A_2		8.2×10^{-4}	2.5×10^{-3}	1.3×10^{-3}	1 : 3 : 1.6
60—65 cm » B		1.3×10^{-3}	3.5×10^{-3}	1.3×10^{-3}	1 : 2.7 : 1
110—115 cm » C		6.7×10^{-4}	2.2×10^{-3}	1.5×10^{-3}	1 : 3 : 2.2
Sérozems (grey-soil) of semi-arid regions	Terek-Kuma semidesert, Bakress Soviet farm				
0—5 cm, hor. A		3.4×10^{-4}	1.0×10^{-3}	5.1×10^{-4}	1 : 3 : 1.5
65—70 cm » B		6.1×10^{-4}	2.8×10^{-3}	8.4×10^{-4}	1 : 4.6 : 1.4
160—170 cm » C		8.8×10^{-4}	2.9×10^{-3}	1.2×10^{-3}	1 : 3.3 : 1.4
Pillar solonetz	Semichnaya station, Rostov region				
0—5 cm, hor. A		8.8×10^{-4}	3.3×10^{-3}	2.1×10^{-3}	1 : 3.8 : 2.4
10—15 cm » B		1.3×10^{-3}	4.8×10^{-3}	2.3×10^{-3}	1 : 3.6 : 1.8
45—50 cm » C		1.3×10^{-3}	4.8×10^{-3}	2.6×10^{-3}	1 : 3.6 : 2
Krasnozems	Batumi Bot. Gardens				
0—10 cm, hor. A		1.0×10^{-3}	2.7×10^{-3}	2.7×10^{-3}	1 : 2.7 : 2.7
40—50 cm » B		1.4×10^{-3}	6.4×10^{-3}	1.1×10^{-2}	1 : 4.6 : 8.0
90—100 cm » C		1.5×10^{-3}	6.5×10^{-3}	1.7×10^{-2}	1 : 4.3 : 11.3
Red-brown clayey steppe soil from ochreous Miocene clays	Shelekta nickel deposit, Aktiubinsk region				
0—5 cm, hor. A_1		3.6×10^{-2}	3.1×10^{-1}	1.3×10^{-2}	1 : 9 : 0.3
45—50 cm » A_2		1.5×10^{-2}	1.3×10^{-1}	1.1×10^{-2}	1 : 9 : 0.7
100—105 cm » B_1		8.5×10^{-3}	6.7×10^{-2}	7.8×10^{-3}	1 : 8 : 0.8

Soils	Place of sampling	Content in % to dry soil			Ratio
		Co	Ni	Cu	Co : Ni : Cu
Brown clayey steppe soil from serpentinites	Novo-Akkerman deposit, Khalilovo region				
0—5 cm, hor. A_1		2.1×10^{-3}	6.2×10^{-1}	2.1×10^{-3}	1 : 30 : 1
25—30 cm » A_2		8.4×10^{-3}	2.0×10^{-1}	1.1×10^{-3}	1 : 24 : 1.3
45—50 cm » B_1		8.6×10^{-3}	1.6×10^{-1}	1.1×10^{-3}	1 : 20 : 1.3
Brown clayey steppe soil	Batamsha nickel deposit, Aktiubinsk region				
0—5 cm, hor. A_1		9.5×10^{-3}	6.2×10^{-2}	7.8×10^{-3}	1 : 6.5 : 0.8
Podzol clayey soil	Tyulenevsky nickel mine, Ufaley deposit				
0—5 cm, hor. A_1		1.2×10^{-3}	1.3×10^{-2}	1.8×10^{-3}	1 : 10 : 1.5
Turf-swamp soil	Ivanovsky iron mine, Alapaevka region				
0—5 cm, hor. A		7.5×10^{-3}	1.0×10^{-1}	1.2×10^{-2}	1 : 13 : 1.6
Podzol loamy soil	2nd Poskotinsky iron mine, Alapaevka region				
hor. A_0		—	1.0×10^{-2}	—	—
0—5 cm, hor. A_1		1.3×10^{-3}	4.6×10^{-3}	3.5×10^{-3}	1 : 3.5 : 3

These data on the cobalt, nickel and copper content of zonal soils of the Soviet Union and those available in the literature on the content of these metals in the soils of other countries permit to form an idea as to the variations and the average content of these metals in soils (clarkes). The average cobalt content of soils is 1.0×10^{-3} per cent; that of nickel, 4.0×10^{-3} per cent; and that of copper, 2.0×10^{-3} per cent. The ratio of clarkes of cobalt, nickel and copper in soils is 1 : 4 : 2, as distinguished from the ratio known for rocks, which is estimated at 1 : 10(15) : 5. The cobalt, nickel and copper content of common soils which are not associated with deposits of metals varies within a narrow range. On the other hand, the soils from the nickel deposits of the South Urals sharply increase in nickel, the Co : Ni ratio attaining 1 : 20.

A certain accumulation of cobalt, nickel and copper in soils rich in humus, *e. g.* in chernozem soils, as borne out in the table, is accounted for by their being invariably present in plants that favour the accumulation and fixing of heavy metals in soils.

When considering the evidence available on cobalt, nickel and copper content of different soils, these may provisorily be divided into such as are rich and such as are poor in those metals. Thus, the podzol, swamp and serozem soils are, as a rule, poor in cobalt, nickel and copper. The deficiency of these metals in the soils is responsible for a number of plant and animal diseases. For example, cereals grown on swampy or highly podzolized soils deficient in heavy metals, copper in particular, are subject to the so-called «reclamation disease». Soils poor in cobalt, *e. g.* the serozems (grey earth) of the arid regions of Australia and New Zealand, and probably of other deserts too, are responsible for a disease of cattle (coast disease), *etc.* An addition of the deficient metals to the animals' diet results in their recovery from the disease. The chernozems are richer in co-

balt, nickel and copper and in other metals (Fe, Ti) as compared to other soils. The presence of appreciable amounts of heavy metals in chernozem soils which represent a soil province enriched with metals, undoubtedly contributes to their exceptional fertility.

An investigation of the Uralian soils sampled over nickel and cobalt deposits revealed there a biogeochemical province highly enriched with cobalt and particularly in nickel. The nickel content of the soil taken over the Novo-Akkerman nickel deposit exceeds 0.5 per cent. It is 155 times the average Ni content of soils. The highest cobalt content amounting to 3.6×10^{-2} per cent has been recorded for soil taken over the Shelekta nickel deposits. The relatively low value of the Co : Ni ratio in this soil, which is equal to 1 : 9, corresponds to the considerable cobalt enrichment of the Shelekta nickel ores. The copper content of the soils taken over nickel deposits is not high; it is either equal to or lower than the cobalt content.

A chemical analysis of the plants collected from the soils overlying nickel deposits showed a very high concentration of nickel in the ash of *Stipa Capillata*, *Artemisia* sp., *etc.* amounting to 6.4×10^{-1} per cent, which is some hundred times the usual nickel content in these soils.

The cobalt, nickel and copper concentration in the soils and plants sampled over the nickel deposits of the South Urals, where the nickel ores generally lie at a small depth (10—20 m), is evidence of a considerable dissemination of these deposits. The determination of nickel, copper and cobalt in soils thus permits to find the areas of dissemination of nickel-cobalt deposits and may serve as a search method for nickel and cobalt ores.

The determination of cobalt, nickel and copper in soil specimens taken along a vertical profile showed their non-uniform distribution in individual horizons, which is an evidence of the presence in the soils of mobile metal combinations, carbonates, humates, *etc.* Nearly in all of the soils, except for the krasnozems (red soils) and serozems (grey soils), there is recorded an accumulation of cobalt, nickel and copper in the upper horizon A and a relative impoverishment of horizon A_2 in the same metals. The Co : Ni: Cu ratio in all the horizons of soils remains approximately constant. An investigation of soil extracts from the medium-podzol soil of the forest plot of the Timiriazev Agricultural Academy revealed, however, a sharp disturbance of the ratio of nickel to copper (there was found Co 3.0×10^{-7} per cent, Ni 1.5×10^{-6} per cent, and Cu 3.1×10^{-6} per cent) towards a relative increase in the copper content. This is evidence of a greater mobility of copper in soils and, hence, of a higher possibility of its being carried with soil and ground waters to rivers and seas.

The distribution of cobalt, nickel and copper in soils depends on the nature of the soil-forming process and on the chemical composition of the rocks. Between the soils and the underlying rocks a secular geochemical relation is established. This relation is brought about by a continuous migration of chemical elements with soil, ground and edge waters and is maintained throughout the geological time. The uppermost film of the Earth's crust, the soil, thus becomes the carrier of many properties, most interesting from the geochemical standpoint, reflecting the distribution of heavy metals in rocks, ores and minerals at depths, hidden from the geologist at the initial stage of his study.

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MINERALOGY

ON THE AGE OF THE INTRUSIONS OF THE NORTH-EASTERN PART OF THE LITTLE CAUCASUS

By A. N. SOLOVKIN

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The question as to the correctness of the age determination of the granodiorite intrusions of the Kedabek, Dashkesan, Chiraghidzor ore deposits, as well as of all the intrusions of the northern slopes of the Shakhdag and Murovdag ranges, comes now to the fore, being intimately related to that of the depths of erosion, the depths of intrusions in general, and to other problems of first rate importance. It was Paffenholz (4) who referred these intrusions to the Oligocene and even to the Lower Miocene, and this view is at present being shared by most of the Caucasian geologists.

The author's works (1-2) showed, on the contrary, that they all should be referred to the Upper Cretaceous cycle: for some of them this age has been proved by the transgressive attitude of Cretaceous rocks overlying them, while others have been correlated to them by analogy. The reasons that permit to assign to all the intrusive massifs, from the so-called Mekhmanin in the South-East and to the Tauz massifs in the North-West, a common age are as follows:

1) All the massifs lie in one geotectonic zone characterized by the same complex of deposits (Jurassic-Cretaceous), the same trend of all the folds and by the same relations to the wall-rocks in general. Typically enough, nearly all of the massifs break mainly Jurassic rocks, being but rarely in contact with Cretaceous rocks.

2) One of the massifs, the Mekhmanin, is transgressively overlain by marls belonging to a substage of the Senonian.

3) The rocks of the massifs are in general characterized by a similar chemical composition, with slight local variations only.

4) Mineralogically all of the massifs are similar, the granodiorite type of rocks being prevalent.

5) The vein series accompanying the massifs is highly characteristic and monotonous, both in orientation (largely of north-western strike) and composition.

Shortly speaking, we have here a combination of geotectonic and petrological features determining one age intrusive series, whose individual elements can well be referred to a definite age on the ground of purely stratigraphical indices.

Recent investigations have furnished additional evidence supporting the author's view. In the Maikop deposits of Naftalan which, as shown by Ali-Zade(4), were being formed, like the Maikop deposits of the area considered here, throughout the Oligocene, and possibly during the Lower Miocene too, Mikhailov(5) discovered feldspars (up to 90 per cent of the fine fraction), fragments of rocks

(quartz-porphyries and other acid effusives), quartz, zircon, garnet, pyroxenes, amphiboles, tourmaline, barite, *etc.*

According to Mikhailov, such a composition of pebbles of the Naftalan conglomerate suggests that the accumulation of sediments in this region proceeded at the expense of an erosion of the rocks of the Little Caucasus and was taking place, as it seems, in the coastal zone of the Maikop basin.

The presence of garnet, tourmaline, pyroxene and feldspars in the sediments of the Maikop series suggests the idea that, among other rocks, the intrusives of the Little Caucasus and their contacts also supplied material for the Maikop series. The presence of barite (!) points to the fact that the hydrothermal veins of barite genetically related to these intrusions also supplied material for the Maikop series. It means that during the Oligocene the erosion had attained such an intensity as to expose the wall-rocks of the intrusives and possibly the intrusives themselves, too.

Thus, it is quite evident that the age of the granodiorite intrusions of the given portion of the Little Caucasus, inclusive of the Dashkesan, Kedabeck and Chiraghidzor intrusions, is much older than the Oligocene and corresponds, as has previously been pointed out by the author (2), to the Upper Cretaceous, within the limits of the Turonian-Campanian.

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GENETICS

TETRAPLOIDY IN CULTIVATED BUCKWHEAT (*FAGOPYRUM ESCULENTUM*)

By V. V. SAHAROV, S. L. FROLOVA and V. V. MANSUROVA

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For about half a century the agricultural science in this country has been facing the problem of how to control the regular reduction in the acreage under buckwheat. In spite of its first-rate economical qualities and its being no exacting crop, neither with regard to soil nor even agrotechnics, buckwheat has a evil repute of a variable and sometimes very poor yielder.

The problem of getting rid of this principal defect of buckwheat can hardly be solved unless the nature of the plant is entirely remodelled. When thinking over this problem, we decided to start by the production of autopolyploid forms. It is to this essentially completed part of our work that the present note is devoted.

The work was started in spring 1941. In our experiments with cultivated buckwheat we seemed to have had no predecessors. In 1930 there appeared, however, a paper on autoploid production in *Fagopyrum tataricum* (1). The writer reported remarkable new characters of the polyploids obtained by him (pronounced increase in size and in seed number per plant).

Our first effort was to obtain as many polyploid plants of different origin as possible. The diversity of varieties and the intravarietal diversity of the hereditary characters was likely to reveal to us the picture of the wide-scale variation of the new forms, involving several characters.

For the production of polyploids we used colchicine in water solution, at concentrations ranging from 0.02 to 0.2 per cent. The treatment was applied to seeds or roots, or else to the point of growth of the stalk.

In all experimental plants there was inhibition of growth, which varied according to the case; at the same time the plants showed typical alterations. These were especially conspicuous in case colchicine was applied to roots.

Some 30—40 days after treatment, when the control plants showed opulent growth, the experimental ones still remained short, stumpy, formed abnormal branches and thick but small leaves. Many of the experimental plants seemed to be arrested in their growth, and a considerable part of them perished. In the subsequent, however, all of the plants that had survived, produced vigorous and flowering shoots. Some of the young shoots stood out for a whole array of characters typical of polyploid plants.

In autumn 1941 a rather considerable amount of seed was collected from the experimental plants, among which we had to sort out the polyploid progeny.

At first we believed cytological examination to be the only method for judging of the polyploidy. The seeds of the experimental plants were examined selectively or in bulk, if the material was no extensive one. With this aim in view the root tips of very young seedlings grown from them were fixed with Carnoy solution, and temporary aceto-carmine preparations were made of them.

Owing to the large number of mitotic plates and the ease of chromosome counts, we usually were able at once to identify the diploid (16), tetraploid (32) or triploid (24) chromosome sets (triploids are very rarely met in our practice).

Table 1

Results of Cytological Examination of Seeds Collected from Colchicine Treated Plants

Variety	Colchicine treatment			Experimental plants		Progeny tested				Seeds tested			
		Dose in %	Time in hours	Total	Gave seed	Only 2n	2n and 4n	Total	% 4n	2n	3n	4n	Total
Bolshevik	Seeds	0.2	48	50	8	7	1	8		52	—	1	53
Bogatyr		0.2	48	50	7	4	1	5		18	—	1	19
Burjat-Mongolskaya		0.2	24	50	6	4	2	6		9	—	3	12
				150	21	15	4	19	21%	79	—	5	84
Bolshevik	Roots	0.05	24	93	47	11	36	47		258	4	162	424
Spanish		0.1	24	49	6	1	5	6		18	—	14	32
Kharkovskaya		0.1	24	62	11	—	10	10		16	1	27	44
Altaiskaya		0.1	24	50	7	2	5	7		19	1	17	37
Belorusskaya		0.1	24	70	8	5	3	8		35	—	4	39
Bogatyr		0.1	23	42	2	1	—	1		2	—	—	2
Kazanskaya		0.1	21	70	5	2	1	3		12	—	1	13
Kazanskaya		0.05	24	93	10	6	4	10		44	—	7	51
				529	96	28	64	92	70%	404	6	232	642
Gomelskaya	Stems	0.1	24	10	2	2	—	2		3	—	—	3
Bogatyr		0.1	27	57	28	23	4	27		175	—	7	182
Bogatyr		0.1	48	14	2	1	1	2		6	—	2	8
T. A. A.*		0.1	24	38	11	10	1	11		29	—	2	31
T. A. A.		0.02	24	36	9	4	5	9		25	—	11	36
				155	52	40	11	51	22%	238		22	260
				834	169	83	79	162		724	6	259	986

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In this way we investigated 986 seeds obtained from 162 experimental plants (Table 1). An analysis of the seeds collected from the initial experimental plants shows that polyploidy can easily be induced in buckwheat plants treated with colchicine. The highest per cent of plants that yielded tetraploid seeds was obtained from those experiments where colchicine was absorbed by the roots directly from solution.

In none of the tests did we succeed to obtain from the initial plant purely polyploid progeny: from each plant that has yielded polyploid seed, normal seeds have also been collected. The fact that the experimental plants sent forth both aberrant and normal sprouts points to the co-existence of diploid and tetraploid tissues.

The almost total lack of triploid seeds shows that, in spite of the presence on the experimental plants of both diploid and tetraploid flowers, re-pollination takes place among flowers of an equal ploidy only. This is of great practi-

cal importance, for tetraploid and control plants can thus be sown on the same plot without running the risk of intermixing.

As a result of all our experiments with the initial plants, polyploid forms in nine bred and indigenous buckwheat varieties have been obtained. Such was the picture revealed to us by the cytological analysis carried out in 1942.

Left, seeds of tetraploids; right, seeds of diploids. Top row, husked seeds; bottom row, unhusked seeds.

We were now able to investigate the properties characterizing the tetraploid buckwheat. It should be noted that before our cytological analysis was achieved, we apprehended the possibility of re-pollination among tetraploid and normal plants, and took therefore care thoroughly to exterminate all plants that yielded diploid sets. This precaution proved unnecessary, and when working with the second generation of tetraploids obtained in the same year directly upon the first one, we started a comparative study of the normal and autopolyploid forms, sowing them side by side.

The polyploids were distinguished by their vigorous development. They may easily be recognized among the rest of buckwheat seedlings by the thicker cotyledon hypocotyl. The cotyledons are also distinguished by greater size and thickness in the 4*n* form. In adult forms these distinctions are still more conspicuous. The tetraploids develop a more vigorous growth, which leads to the formation of a thick stalk and of vigorous branches. The leaves on them are much larger, their colour is deeper. The surface of the leaf-blade is uneven, the blade itself is thick and coarse.

Table 2

Average Seed Weight in Normal (2*n*) and Autotetraploid (4*n*) Forms

Variety	Grain weight in mg		Weight in 4*n* forms in %
	2*n*	4*n*	
T. A. A.	26.6	38.82	146
Bolshevik	25.37	37.01	146
Bogatyr	24.0	39.59	165
Kharkovskaya	23.56	33.46	142
Spanish	23.4	38.05	163
Burjat-Mongolskaya	21.56	39.91	185
Kazanskaya	21.05	32.39	154
Altaiskaya	19.7	35.57	181

The polyploids start flowering about the same time as the normal plants, but their blossoms are usually much bigger, all the parts of the blossom being correspondingly increased. Of interest are measurements of the pollen-grains. Buckwheat is a heterostylous plant; it has been known long since that the pollen of its short-style flowers is coarser than that of the long-style ones. In our material the polyploid forms of the flowers of both categories gave pollen whose size was so much above the normal that it was sufficient to measure a dozen of pollen grains to safely determine whether a given plant was a polyploid or a normal form. We thus find ourselves in possession of another, extremely simple procedure to determine polyploidy in a flowering plant.

In any plant of cultivated buckwheat polyploidy leads to a conspicuous increase in the volume and weight of the seed (Table 2 and figure).

This increase is especially pronounced in the eastern indigenous varieties of the Burjat-Mongol and Altay buckwheat which, along with fertility, are known to give small seed inconvenient for husking.

The increase in the total weight of polyploid seeds (Table 3) is but partly dependent on the relatively greater mass of husk in these seeds: the mean weight of husked seeds exceeds that of husked diploid seeds by $^1/_3$. It should, however, be pointed out that the per cent of the weight of the husk, as borne out on our material, is as variable a character as most of the characters of different polyploid forms. No correlation between the per cent of husk and the increase in the weight of the seed could be established, the coefficient being—0.12. Another important question was to be cleared, *viz.* is there not a reverse relation between the number of seeds per plant and their weight? It is easily understood that when small in number, the seeds of a plant will enjoy better food and ripening conditions. But no correlation could be found between the number of seeds per plant and the mean weight of a single seed, the coefficient being—0.11, either.

Table 3

Average Data and Seed Weight in Diploid and Autotetraploid Forms (all the sorts together)

Average weight of non-husked seeds in mg	$2n$	24.65 ± 0.76	$\sigma = \pm 3.96$
	$4n$	36.71 ± 0.36	$\sigma = \pm 4.95$
Average weight of husked seeds in mg	$2n$	21.56 ± 0.57	$\sigma = \pm 2.98$
	$4n$	28.64 ± 0.37	$\sigma = \pm 5.01$
Per cent weight of husk to weight of non-husked seeds	$2n$	17.22 ± 0.73	$\sigma = \pm 3.81$
	$4n$	26.42 ± 0.36	$\sigma = \pm 4.92$

From a comparison of the 2nd generation of the tetraploids which achieved their developmental cycle by October 1942, with the control diploid plants, the former were found to have achieved their growing period on the average within 107 days, while the latter did so within 102 days. But the tetraploids were less uniform than the normal forms. It should be noted that by the length of their growing period and other characters our new plants widely differed from one another even within a single variety.

Highly instructive in this respect are data secured for the most important character, the fertility, *i. e.* number of seeds that ripened on each separate plant. Along with plants that yielded an insignificant amount of seed, there were such as gave more than a hundred, and in three cases, even more than two hundred seeds*.

At any rate we have no ground whatever to speak of a reduced fertility in polyploids. The value of our material consists in that it permits of a comparative study of polyploid forms differing as to their fertility. Work along this line may help to reveal the cause of the reduced fertility usually displayed by new experimentally created polyploid forms of most plants.

It remains to point out the wide range of variation observed among our new forms. The large number of most diverse characters, involving size, branching, flowering, shape and colour of seed *etc.*, recorded by us, undoubtedly represents a most valuable material for the breeding of buckwheat. At the time being polyploid buckwheat offers new forms, with large seed, readily husked because of its loosely adhering husk. But the extreme variability of these forms is greatly promising on account of a whole set of other economically valuable characters. This is why, even without interspecific or intervarietal hybridization, the polyploidy in buckwheat opens wide prospect for the breeding of highly productive and resistant forms of this crop.

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* In summer 1943 we collected a crop from the third generation of tetraploids. By having selected for propagation the most fertile, coarse-grained forms, we have succeeded at the very start of our breeding work to attain a sharp increase in yield.

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MICROBIOLOGY

PREPARATION AND PROPERTIES OF CRYSTALLINE GRAMICIDINE C

By G. F. GAUSE, M. G. BRAZHNIKOVA and N. P. LISSOVSKAYA

(Communicated by L. A Orbeli, Member of the Academy; 24.XII.1943)

The American microbiologist Dubos (1) isolated in 1939 from a culture of a particular strain of an aerobic sporulating bacillus a powerful antibacterial substance. The substance kills staphylococci, streptococci and pneumococci, without causing any harm to the infected tissues of the body.

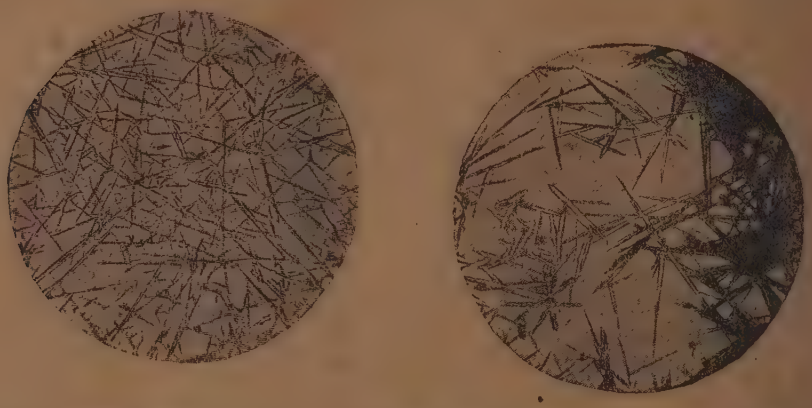
In 1941 Dubos and Hotchkiss (2) found that this substance, henceforth called tyrothricine, represents a polypeptide consisting in two crystalline fractions: gramicidine (15 per cent) and tyrocidine (about 85 per cent by weight). Subsequent experimental and clinical work has shown tyrothricine to possess considerable therapeutic value, so that the latter is now being used for the healing of infected wounds. Judging by the strength of the antibacterial action of the crystalline fractions, gramicidine appears to be the most valuable component of tyrothricine. However, it is not used widely in hospitals of the USA because of its small yields (about 40 mg per litre of bacterial culture), and of the considerable difficulties of preparing it on a commercial scale.

In the present work it is reported that practically pure gramicidine (about 95 per cent by weight of the antibacterial preparation) is formed by a new strain of bacilli isolated by Gause and Brazhnikova (3) from Moscow soils. This strain yields about 300 mg of crystalline gramicidine per litre of the bacterial culture, under mass production. In view of the absence of the tyrocidine fraction and certain differences between our preparation and that of Dubos (our preparation is soluble in chloroform), the crystalline substance obtained by us will henceforth be called gramicidine C. Gramicidine C has already shown its high therapeutic value in a number of hospitals, and the strain of Gause-Brazhnikova is now being used for the preparation of this substance in bulk.

Preparation of Gramicidine C. Bacteria are grown in a 10 per cent yeast autholysate for 6 days, the culture fluid is acidified by muriatic acid to pH = 4.7, the sediment filtered and thrice extracted by alcohol. The latter is concentrated *in vacuo* to 1/10 of the volume and then diluted 10 times by an 1 per cent solution of NaCl. Gramicidine precipitates, after which it is dried, dissolved in alcohol and sealed in ampulae in such a way that 1 ampula contains 20 mg of gramicidine per 0.5 cm^3 of alcohol. Before using, the contents of an ampula is diluted 50 to 100 times with sterile water, and the solution of gramicidine is applied in tampons for treating infected wounds, or for washing fresh wounds with the purpose of prophylaxis.

Chemical Properties of Gramicidine C. The dry preparation of gramicidine is washed with petroleum ether for the removal of lipoids. Gramicidine is insoluble in this solvent. It represents a gray powder containing 13% per cent of nitrogen (Kjeldahl). It is readily soluble in alcohol

and acetone and crystallizes from them in the form of platelets with pointed ends. It is also soluble in acetic acid. These properties of our preparation are identical with those of crystalline gramicidine of Hotchkiss and Dubos. The fact that our preparation consists of the gramicidine fraction alone and contains no tyrocidine fraction is seen from the following observations. When the initial preparation of Hotchkiss and Dubos is being extracted by a mixture of sulphuric ether with acetone (1 : 1), only 15 per cent by weight of it dissolve (the gramicidine fraction), whereas 85 per cent remain insoluble (tyrocidine). The latter does not crystallize from neutral solutions and cannot be obtained in the crystalline form unless muriatic acid is added. It crystallizes as tyrocidine chlorhydrate. Unlike this, 95 per cent of our initial preparation is soluble in a mixture of sulphuric ether with acetone, and is further wholly crystallizable from neutral alcoholic or acetonic solutions. It should be noticed that the crystalline gramicidine



Crystals of gramicidine C from alcohol (to the left) and from acetone (to the right).

described in this work differs from the crystalline preparation of Hotchkiss and Dubos by its solubility in chloroform. Further chemical study of gramicidine C has to reveal the details of its molecular structure.

Action of Gramicidine C upon Pathogenic Bacteria. In the testing of the effect of gramicidine upon various strains of bacteria the Metchnikoff Institute and the Tarassevitch Control Institute took part. The nutrient broth containing gramicidine at increasing concentrations, from 1 to 400 gamma per 1 cm^3 of medium (gamma=1/1000 mg) was inoculated with a fresh culture of bacteria. The test tubes were incubated at 37° C for 24 hours, and by the presence or absence at that time of bacterial growth the degree of bacteriostatic efficiency was evaluated. The bactericidal effect of gramicidine was judged upon from the results of planting the contents of the tubes upon the solid media (table).

Toxicity of Gramicidine C. The average toxic dose of gramicidine for white rats, killing 50 per cent of the animals, makes up about 15 mg per kilo weight for intraperitoneal injections. Hotchkiss' and Dubo's gramicidine possesses a similar toxicity. When applied *per os* gramicidine C is harmless at a dose of 100 mg per kilo weights, and solutions containing 400 gamma of gramicidine per 1 cm^3 can safely be injected into the pleural or peritoneal cavity, at a dose of 3 mg per kilo weight. Mice survive subcutaneous injections of 1 cm^3 of a solution containing 800 gamma of gramicidine. The local application of gramicidine solution containing 400 and 800 gamma of this substance per 1 cm^3 kills pathogenic cocci and anaerobic bacteria without injuring the tissues of the infected animal.

Prophylactic Action of Gramicidine C. The gramicidine prophylaxis against experimental gas gangrene in guinea-pigs was studied

Minimum Bacteriostatic and Bactericidal Concentrations of Gramicidine C, Expressed in Gamma, for Various Bacteria

Microbes	Concentration of gramicidine C	
	bacteriostatic	bactericidal
Staphylococci	6—12	25—50
Streptococci	6—12	12—25
Pneumococci	100	100
Gonococci	100	400
Cl. Welchii	10—25	10—25
Cl. histolyticum	10—25	10—25
B. diphteriae	50	50
B. anthrax	10—50	Spores are not killed
B. coli	No action	—

by N. V. Zhuravskaya at the Metchnikoff Institute. In 33 animals the muscle tissue of the hip was severely injured, and infected with one-milliard suspension of *Cl. Welchii*. Out of 15 control animals 8 died, and 7 displayed heavy necrosis of the muscular tissue. Among 18 animals treated by local application of gramicidine solution, 11 remained healthy, 1 died, and 6 showed only slight necrosis of the muscular tissue.

In order to approach more closely the actual conditions of war surgery, we infected with earth the injured muscles of the hip in white rats. In the control animals the wounds were infected with earth, covered by gauze moistened with saline solution and tightly sutured. In the experimental animals the gauze was moistened with a solution of gramicidine containing 800 gamma of this substance per 1 cm^3. Forty-eight rats were kept under observation. In one set rats weighing 200 g were infected with 25 mg of dry earth each. Half of the control animals died from anaerobic infection, whereas all the animals treated with gramicidine remained alive. In another set, young rats weighing 100 g each were used. The mortality in the control rats infected with the same dose of earth attained 100 per cent, whereas the mortality among animals treated with gramicidine was as low as 40 per cent. These experiments prove the prophylactic efficiency of gramicidine against anaerobic infection.

Clinical Observations. The treatment with gramicidine of 100 clinical cases, carried on under the supervision of Prof. N. N. Priorov and Prof. I. G. Rufanov has shown the efficiency of this substance. The wounds with long-lasting infection got rapidly free of pathogenic cocci, body temperature dropped to normal values, and the wounds rapidly granulated with healthy epithelization from the edges. Cytological observations failed to reveal any harmful effect of gramicidine upon the regenerating surface of the wound. Acute traumas due to street accidents treated with gramicidine solution healed in all cases without suppuration.

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and Medical Parasitology.

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PLANT PHYSIOLOGY

ON THE CULTURE OF SOME MEDICINAL PLANTS IN THE NOVOSIBIRSK PROVINCE

By W. W. REVERDATTO and A. G. GONCHAROV

(Communicated by A. A. Richter, Member of the Academy, 8.XII.1943)

I. *Lobelia inflata*

Under the war conditions we found ourselves faced by the problem whether some of the important medicinal crops can possibly be transferred from the Ukraine and the Northern Caucasus, and from the European part of the USSR as a whole, to Siberia, in particular, to the Novosibirsk province. One of these crops was the well-known medicinal plant *Lobelia inflata*. Preparations of *Lobelia* are of special importance as a means against gas poisoning. Before the war, *Lobelia* was being successfully cultivated in the western regions of this country, in White Russia, as well as in the Moscow and Ivanov provinces. In the vicinity of Leningrad it gave rather poor crops, though the content of the alkaloids attained 0.40 per cent. In 1941, in the vicinity of Tomsk, *Lobelia* plants grown from seedlings developed a good green mass, flowered and gave capsules, but seed failed to ripen on the plot; part of the plants have therefore been transplanted into pots, and the seeds made after-ripening under greenhouse conditions.

In 1942 we were put before the following task: to make the seed ripen in the ground, as well as to find out the optimal conditions for obtaining abundant yields of green mass along with high contents of the alkaloids.

In order to solve the former problem experiments were carried out in which the plants were exposed to the effect of a curtailed day. With this purpose in view young seedlings planted out in boxes were treated, from the moment of sprouting and until they were planted into soil, with short-day illumination. This was attained by screening them since 8 p. m. until 8 a. m. and thus creating for them a 12-hour day. The control plants received a normal 15—18-hour illumination. As soon as after a month the exprimental plants formed 4—5 leaflets, whereas the controls had no more than 1—2 leaflets each. The short-day plants started flowering on July 25; the controls, on August 10.

The average height of a short-day plant was 50—60 cm; that of controls, 30—40 cm.

Seeds were harvested in short-day plants by September 10, the yield collected from a plot of 100 m² being 600 g of ripe seed. The seeds of the controls were not ripe yet. Thus, the effect of light treatment proved a positive one.

The second experiment was carried out in order to estimate the effect of various fertilizers upon the yield of dry mass and content of alkaloids. Each plot measured 100 m². Harvesting was done on September 4, plants air-dried.

The yield of dry mass in tons per 1 ha was as follows.

Table 1

Controls	1.15
Short-day plants	3.85
With P_2O_5	4.15
With K_2O	3.0
With N	2.2
With $P_2O_5 + K_2O + N$	3.8

Samples of dry mass were taken from each experimental plot for analysis. At the laboratory of plant physiology and biochemistry of the Tomsk State University the material has been analysed to determine the amount of alkaloids contained in *Lobelia inflata*; the analysis was made under the direction of Prof. Sukhorukov, after the gravimetrical method of Bertrand and Javellier. The results obtained are tabulated below.

Table 2

	Alkaloid content in %
Control	0.3
Short-day plants	0.3
With P_2O_5	0.33
With K_2O	0.45
With N	0.34
With $P_2O_5 + K_2O + N$	0.24

II. *Atropa Belladonna*

The problem of cultivating *Atropa Belladonna* under the conditions prevailing in the Novosibirsk province presents considerable interest, for up to recently this plant was cultivated mainly in the European part of this country, and in its southern regions with that.

Attempts to naturalize this plant in the Altay region were made in 1942 and failed for several reasons.

Experiments with *Belladonna* were started in 1942. Seedlings were planted on an open plot. Towards the fall the plants formed but their near-ground rosettes, and during winter 100 per cent of them were frost-bitten, possibly because of insufficient snow-cover.

In 1943 the experiments were resumed. The aim was now to make the plants develop an opulent green mass, and to obtain seed during the first year of cultivation from plants grown from seedlings and transplanted into soil. The method applied was the stratification of seeds and short-day treatment of seedlings.

The seeds were stratified since autumn into moist sand, and kept in a cellar at a temperature of 1—2° C.

On March 10, 1943, the seeds were planted into boxes and transferred to a green-house. By April 8 they sprouted at one. After this some of the plants were given curtailed 12-hour illumination, from 8 a. m. until 8 p. m., the controls receive dnatural light during about 15—18 hours. The plants were kept so until planting into soil was done.

On June 4, the seedlings were planted into loamy soil; the plot was chosen in an orchard, between trees.

On August 5, the short-day plants gave abundant flowering, while only 25 per cent of the controls bore isolated blossoms on August 25. On August 26 the experimental plants from a plot 70 m² large and the controls were subjected to comparative measurements. The results are tabulated below.

Table 3

	Short-day plants	Controls	Difference
Height of plant	110 cm	50 cm	+ 60 cm
Stalk thickness near root	70 mm	50 mm	+ 20 mm
Number of stalk branchings	22	3	—
Number of large leaves	74	14	+ 60
Number of medium and small leaves	228	51	+177
Average number of flowers per plant	9	6	—
Number of fruit sets	13	3	+ 10

The leaves were collected at the period of flowering, both from experimental and control plants. Gravimetrical estimation gave the following results:

Table 4

Yield in kg per ha	Short-day plants	Controls	Difference	Increment in %
Leaves	660	300	360	220
Stalk	475	150	325	—
Root	310	220	90	—
Total	1445	670	775	215

Chemical analysis of leaves made in the Laboratory of Prof. Sukhorukov, under his personal guidance, has shown that the content of alkaloids in the leaves of experimental plants was 0.37 per cent, while in controls it was 0.34 per cent. The content of alkaloids in the roots was about the same, *viz.* 0.42 and 0.41 per cent, in experiments and controls, respectively. As known the content of alkaloids in leaf, as stipulated by pharmacopoeia, is 0.35 per cent. The fruits of *Belladonna* started ripening on experimental plants by September 5, and became fully ripe in rooms. The plants showed a curious response to the early autumn frosts: the first frosts produced no effect at all, and it was not until the temperature dropped to—4° that the plants perished.

Conclusions

1. *Lobelia inflata* can well be cultivated in the vicinity of Tomsk and in the central belt of the Novosibirsk province; planting should be made by seedlings grown from seeds sown into boxes or hot-beds early in April.

2. By exposing the plants to the action of a short, 12-hour-day, prior to their planting into ground, it is possible to obtain fully ripe seeds.

3. The best results, with respect to both amount of vegetative mass and alkaloid content, are obtained with cultures grown under application of potassium and phosphorus fertilizers.

4. *Atropa Belladonna* can be cultivated under conditions prevailing in the central regions of Siberia as an annual plant; it should be grown on well protected plots or among the trees of an orchard, or on forest glades.

5. When pre-treatment with a short, 12-hour-day is made to precede planting into ground, the seedlings of *Belladonna* give a better yield of the useful product; the content of the alkaloids is somewhat increased thereby. The treatment makes it possible to collect fruits which readily attain full ripeness when stored in a room, and give germinable seed.

6. The problem of an annual culture of *Atropa Belladonna* in Siberia, both for leaf and for seed, may thus be considered solved.

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PHARMACOLOGY

ON THE QUESTION OF THE BIOLOGICAL ACTIVITY OF ABOVEGROUND PARTS AND ROOTS OF *ADONIS TURKESTANICUS*

By N. KLEMPARSKY

(Communicated by E. N. Pavlovsky, Member of the Academy, 7.XII.1943)

In our investigation of the biological activity of *Adonis turkestanicus* wild growing in Central Asia, we have scarcely had any predecessor. The only pharmacological study devoted to this species belongs to Kamilov ([8]) who tested the effect of the aboveground parts of this plant on the heart of frog, and found it equivalent to that of the official *Adonis vernalis*.

Of especial interest is the standardization of the roots of *Adonis vernalis*, for they are not being used in modern medicine. Information as to their having been used in old scientific medicine as a cardiac and diuretic is, however, available in literature ([4], [6]).

The plant *Adonis turkestanicus* is widely distributed in the Uzbek and Tadjik SSR in mountainous regions, foothills and along the upper reaches of mountain rivers there([2], [7]).

Vegetable material for our investigation (aboveground parts in the form of stalks, leaves, unripe fruits, as well as roots of *Adonis turkestanicus*) was obtained early in March 1942 from the Botanical Institute of the Tadjick Branch of the Academy of Sciences of the USSR.

This material was collected in August 1941 on the southern slopes of the Ghissar mountain range, along the upper reach of the river Kapandar, *i. e.* within the zone of sub-alpine vegetation. The material has well preserved its original aspect; colour and smell. No signs of deterioration could be observed.

The root material consisted of rhizomes with stalks, and of the root proper; it was properly dried, there was some soil sticking to it, as well as root hairs of some other plants. The roots were 5—10 cm long. Their surface was covered with tubercles; they were spirally wound along the longitudinal axis, bore traces of old detached stalks and were coated with a dark-brown, almost black crust. They were white on cross-section, the inner part very hard, bitter, without any smell.

The aboveground parts and roots of *Adonis turkestanicus* were tested for their biological activity in strict conformity with the main requirements of the USSR Pharmacopoeia, Edition VII ([1]) and with commentaries to the latter to be found in a book by Sarguin ([5]). Alcohol extracts were prepared by U. P. Kalashnikov from ground material. Each dose of the extract was tested on 5 specimens of *Rana ridibunda*, records being taken of the systolic stop of the ventricle by the end of the first hour (or 1 hour and 10 min. at most) in all of the frogs or in most of them (in 3—4 out of 5). The experiments were run at a temperature of 20—25°C.

Previous to injection the extracts were diluted fourfold with distilled water. If upon dilution the extracts became appreciably turbid, they were added 1—2

drops of one per cent solution of sodium carbonate, after which the liquid got clear rapidly and was at once introduced by means of a thin needle under the skin of one of the thighs; if the dose exceeded 0.3 ml, under the skin of both thighs in equal amounts.

The lowest dose of alcohol extract from aboveground parts of *Adonis turkestanicus* that was found to cause a full and irreversible stop of the heart ventricle in all frogs by the end of the first hour (+10 min.) was 0.4 ml, and for the majority of animals it was 0.3 ml. A dose of 0.45 caused a stop within less than an hour.

For lack of a standard extract from the aboveground part of *Adonis vernalis*, the value of the Turkestanian plant has been calculated by us on the ground of our own data, and compared with the requirements of the Pharmacopoeia actually adopted in the USSR. From these data it is evident that one efficiency unit (e. u.) is contained in 0.3—0.4 ml of alcohol extract. Consequently, one gram of the latter contains 2.5 efficiency units.

When recalculated to dry matter of *Adonis turkestanicus* collected in 1941, one gram of it should contain 50—60 e. u., which is in full agreement with the requirements of Pharmacopoeia concerning the evaluation of the effect of *Adonis vernalis* [1].

Experiments on root extracts of *Adonis turkestanicus* showed that 0.2 ml stop the heart of a frog within less than one hour, and 0.1 ml, by the end of the first hour. A dose of 0.05 ml produces no such effect even within one hour or 1 hour and 15 minutes. This should be taken to mean that such an amount is insufficient to estimate the biological activity of the drug.

Thus, 0.1 ml should be considered the least dose; according to its activity, this dose equals 0.3—0.4 ml of the extract from the aboveground parts of *Adonis turkestanicus*; in other words, the effect of the root is found to be three-four times that of the aboveground parts of this plant.

The value (efficiency) of *Adonis turkestanicus* should be estimated from a dose of 0.1 ml. This dose contains one e. u. and 1 g of the extract contains 10 e.u.

If this calculation is applied to pure dry weight of the root, one gram of the latter is found to contain 200 e. u. This figure is almost four times that required by the Pharmacopoeia of the USSR from *Adonis vernalis*.

Conclusions

1. The value of the aboveground parts of *Adonis turkestanicus* should be considered equal to 50—60 frog e. u., which is in agreement with the requirements of the Pharmacopoeia of the USSR for *Adonis vernalis*.

2. The root of *A. turkestanicus* is biologically much more active than the aboveground parts of this plant. Its value equals 200 e. u.

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ZOOLOGY

NEW SUBSPECIES OF BIRDS FROM THE WRANGEL ISLAND

By L. A. PORTENKO

(Communicated by E. N. Pavlovsky, Member of the Academy, 9.X.1943)

A taxonomic treatment of birds collected by me during my stay at the Wrangel Island as a member of the staff of the expedition of the Academy of Sciences of the USSR in 1938—39 enabled me to describe two new forms from this island. The present article contains a description of new subspecies of *Uria lomvia* (L.) and *Cepphus grylle* (L.), along with a revision of all the subspecies of those birds occurring in this country. As material use was made in the first line of the series collected by myself on the Wrangel Island, Chukchee Peninsula and Nowa Zemlya, as well as of the extensive series from the Collections of the Zoological Institute of the Academy of Sciences of the USSR and of the Institute of Arctic Agriculture in Leningrad.

1. *Uria lomvia* (L.), Brunich's Guillemot. Within the boundaries of the USSR this species forms the following subspecies:

1) *lomvia* (L.). Intermediate between the darkest and the palest subspecies. The secondaries broadly tripped with white. Wing 22.1—20.2 (on the average 21.3) cm in males, 22.4—20.7 (21.2) cm in females. 31 specimens examined.

Distribution. I have studied specimens from Murman coast, Nowa Zemlya and Spitzbergen. The subspecies seems to be widely distributed further westward. The only specimen from North-East America I happened to handle (a male from St. Lawrence gulf) seemed to me undistinguishable from our typical *lomviae*; yet, being killed in 1860, it might well have lost the tinge that is peculiar of fresh skins.

2) *arroides* Port. This subspecies was described by me in 1937 [1] after five specimens. Having resumed the treatment in 1940, I disposed already of a small but well prepared series consisting of six specimens. By the proportions of its bill and length of the wing, the new subspecies is similar to the typical form, save for a darker black-blue tint of the upper side, approaching in this character the pacific subspecies, *arra* (Pall.), on account of which it has been named by me *arroides*. White trim on the wing narrower than in the typical form. Wing 22.2—20.1 (21.2) cm in males, 21.8—21.8 (21.8) in females. Four specimens examined.

Distribution. Franz Josef Land.

3) *eleonorae* Port. Described by me along with preceding form [1]. Differs sharply both from the typical form and from *arroides* by its paler colour. The upper side pale-grey with a slate or bluish tint. The coffee-brown of the foreneck paler than in all races, save the next. The streaks on the sides of the body absent or but slightly pronounced, greyish in colour. White tips of secondaries broader than in the typical form. The smaller under wing-coverts white without any brown markings. Bill shorter than in subspecies next described: 2.73—2.58

(2.66) cm from the nostrils, 3.51—3.43 (3.47) cm from forehead in males; 2.58—2.47 (2.53) cm and 3.35—3.22 (3.29) cm in females, respectively. Wing 22.3—21.5 (21.9) cm in males, 21.7—21.5 (21.6) cm in females. Seven specimens examined.

Distribution. The eastern part of Taimyr Peninsula and the New Siberian Archipelago.

4) *heckeri* subspecies nova. Similar to *eleonorae* Port. as to its pale coloration, but differing from the latter in having a stronger bill and a longer wing. In spring plumage the grey colour is of a colder tint. Bill 3.17—2.77 (2.98) cm from the nostrils, 4.00—3.48 (3.79) cm from forehead in males; 3.24—2.67 (2.87) cm and 4.15—3.39 (3.64) respectively in females. Wing 23.4—21.3 (22.2) cm in males, 24.0—20.4 (22.1) cm in females, 46 specimens examined.

Type at the Zoological Institute of the Academy of Sciences of the USSR: ♂, Cape Waring, on the north-eastern coast of Wrangel Island, May 14, 1939. Paratype: ♀, same locality, same date. Both obtained by me. Type-locality: Wrangel Island.

Colour of the upper body side as pale as in *eleonorae*. It is much alike the light colour of the upper side in *Uria aalge inornata* Salom., which should be borne in mind, say, in the Gulf of Providence, where one is likely to meet *Uria lomvia heckeri*, *U. l. eleonorae* and *U. aalge inornata* at the period when they are not breeding. From *eleonorae* the new subspecies is distinguished by its coarser and larger bill, which is more illustrative when directly comparing specimens than in the case of measurements. I took the length of bill from the notch near the nostrils where the lateral prominence of the plumage is ending. These measurements proved of a greater diagnostic value than those along the chord from the limit of the forehead plumage, which are liable to considerable variation. The series of birds collected in May on Wrangel Island in fresh plumage after moult are remarkable for a colder pure blue tint of their grey coloration. Wing longer. Consequently, this form approaches *arra* as to dimensions.

Distribution. Breeding on Wrangel Island and the northern coast of Chukchee Peninsula. Migrates to the Bering Sea. The specimens from Wrangel Island show more conspicuous characters than those from Koluchi Island. The only specimen from the north-western Alaska, a male from Norton Bay, obtained on July 18, 1843, resembles the form described, except for being somewhat darker. At any rate it cannot possibly be referred to *arra*. I gave it its name in honour of Dr. R. Th. Hecker, chief of the Wrangel Island expedition of the Academy of Sciences of the USSR, to whom I am greatly indebted for his constant help in my work.

Table 1

	Length of bill, in cm							
	From nostril				From forehead			
	maxim.	minim.	mean	measur. number	maxim.	minim.	mean	measur number
Males								
eleonorae Port.	2.73	2.58	2.66	3	3.51	3.43	3.47	3
heckeri Port.	3.17	2.77	2.98	22	4.00	3.48	3.79	22
Females								
eleonorae Port.	2.58	2.47	2.53	3	3.35	3.22	3.29	3
heckeri Port.	3.24	2.67	2.87	19	4.15	3.39	3.64	19

5) *arra* (Pall.). Much darker as compared to all other races. The coffee-brown of the fore-neck deeper. The streaks on the sides of the body larger and richer in colour. Secondaries with narrow white trim. Smaller under wing-coverts

marked with brown. Wing 23.2—21.3 (22.4) cm in males, 22.6—20.9 (22.0) cm in females. 38 specimens examined.

In the forms from the Comadore Islands the characters typical of *arra* are more clearly pronounced than in specimens from the southern coast of Chukchee Peninsula. No safe identity with the forms from the Kuriles, Kamchatka, and with the American forms from South Alaska could be established for lack of material.

Distribution. Breeds along the coasts of the Bering Sea, from Bering Strait and further south, along the eastern coast of Kamchatka and the Kurile Islands.

Table 2

Subspecies	Length of wing, in cm							
	Males				Females			
	maxim.	minim.	mean	measur. number	maxim.	minim.	mean.	measur. number
lomvia (L.)	22.1	20.2	21.3	19	22.4	20.7	21.2	12
arroides Port.	22.2	20.1	21.3	2	21.8	21.8	21.8	2
eleonorae Port.	22.3	21.5	21.9	4	21.7	21.5	21.6	3
heckeri Port.	23.4	21.3	22.2	27	24.0	20.4	22.1	19
arra (Pall.)	23.2	21.3	22.4	19	22.6	20.9	22.0	19

II. *Cepphus grylle* (L.), Black Guillemot. The following subspecies are encountered in this country.

1) *grylle* (L.). The typical form is well distinguished by the white wing-patch being cut in two by a black transversal bar. As to colour and wing length, it does evidently not differ from the next subspecies described, save for some specimens of *grylle* having a bill thicker at the base, and of a paler green.

Distribution. Breeds in the North Atlantic, from Ireland, Newfoundland and Labrador up to Southern Greenland and the Murman coast, penetrating also in the southern parts of the Baltic Sea. According to Harter, forms transitional to the next subspecies occur in Greenland.

2) *mandti* (Licht.). This subspecies is distinguished from the preceding one by having no bar across the white wing-patch. The differences in the shape of bill and the green tint of its black plumage seem hardly to be significant.

Distribution. Breeds on either shore of the Polar basin from Greenland, Spitzbergen, eastward through Nowa Zemlya, Taimyr, the New Siberian Islands, along the north-western Chukotka up to Arctic Canada, inclusive of Wrangel Island.

3) *tajani* subspecies nova. Like *mandti* (Licht.) but differing from the latter in its deep sooty-black colour with a very intense greenish gloss. Any specimen of this race may easily be distinguished by its colour. Wing 18.1—16.6 (17.2) cm. 41 specimens examined.

Type in the Zoological Institute of the Academy of Sciences of the USSR: ♂, Cape Pillar, eastern coast of Wrangel Island. Paratype: ♀, same locality, same date. Both collected by me. Type-locality: Wrangel Island.

Distribution. Seems to breed exclusively on the Wrangel Island, perhaps on Herald Island.

Named by me in honour of the eskimo Tayan, former chief of Wrangel Island.

The study of the subspecies of such typical Arctic birds as are Brunich and Black Guillemots permits of the following general conclusions. The divergence of species inhabiting high altitudes into small geographical races is as common a phenomenon as at moderate latitudes. The examples studied by us are but particular cases illustrating the general rule, *viz.* that the formation of geographical races is largely due to geographical isolation. The mode of life of the two species,

Table 3

Subspecies of *Cepphus grylle* (L.)	Length of wing, in cm			
	maxim.	minim.	mean	measur. number
grylle (L.)	17.7	15.2	16.7	25
mandti (Licht.)	17.4	15.0	16.6	40
tajani Port.	18.1	16.6	17.2	41

along with the peculiar conditions of breeding upon rocky shores with cornices and crevisses, were highly favourable for the isolation of subspecies. In order that stable and sharply differentiated types be formed, isolation must last for a sufficiently long time. Both Brunich and Black Guillemots belong undoubtedly to the most ancient inhabitants of the Arctic. Being island and rock dwellers, they survived without any important migration all the events of glaciation and sea-transgression. This is why the range of subspecific variation in them is not narrower than in the inhabitants of moderate latitudes. The high latitude and the special conditions of the Arctic climate do not prevent the species from segregating into indigenous races. The influence of the North has been reduced to a formation of such peculiar distinctive characters as a larger size, a lighter colour and some other traits that usually distinguish northern subspecies.

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